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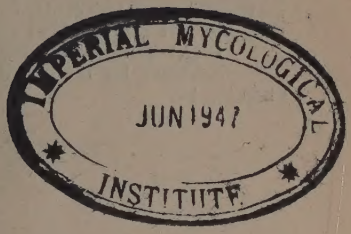
(DOKLADY)

DE L'ACADÉMIE DES SCIENCES DE L'URSS

NOUVELLE SÉRIE

1945



VOLUME XLVII

№ 7

ÉDITION DE L'ACADÉMIE DES SCIENCES DE L'URSS
MOSCOU

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Édition de l'Académie des Sciences de l'URSS
Moscou 1945

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

MATHEMATICS

ON A PROBLEM CONNECTED WITH PURELY DISCONTINUOUS RANDOM PROCESSES

By V. M. DUBROVSKY

(Communicated by A. N. Kolmogoroff, Member of the Academy, 8.VI.1944)

The problem considered here was proposed to the author by A. N. Kolmogoroff.

Let $\mathfrak{A}$ be the set of states of a system $\mathfrak{N}$. Suppose a familly $\mathfrak{M}$ of subsets of $\mathfrak{A}$ is defined, that contains all differences and finite or enumerable sums of its elements, as well as $\mathfrak{A}$ itself, the single elements of $\mathfrak{A}$ and the void set. Suppose the process of evolution of the system $\mathfrak{N}$ has the property that every bounded time interval $t < s \leqslant \tau$ can be divided into a finite (though arbitrarily great) number of subintervals, in none of which the state of the system is altered. Let $\overline{p}(t, x, \tau)$ denote the probability that the state of the system $\mathfrak{N}$, being $x \in \mathfrak{A}$ at moment t, will be altered once during the time $t \leqslant s < \tau$.

Let further $p(t, x, E)$ denote the probability that the state of the system $\mathfrak{N}$ being x at moment t and undergoing an alteration at the same moment, remains in a set E during a sufficiently small time interval $t < s \leqslant t+h$. We assume that $\overline{p}(t, x, \tau)$ is determined by the equality

$$\overline{p}(t, x, \tau) = p(t, x)(\tau - t) + (\tau - t) = \varepsilon \qquad (1)$$

where the function $p(t, x)$ is integrable (in the sense of Riemann) with respect to t, measurable with respect to x and $\mathfrak{M}$, non-negative and less than a constant K; ε becomes infinitesimal as $\tau \to t$, and is less than the same constant K. As to the function $P(t, x, E)$ it is supposed to satisfy the first three conditions imposed on $p(t, x)$ and, moreover, to be completely additive with respect to E and not to exceed 1. All these conditions are introduced for a domain $-T < t < \tau < T$, $x \subset \mathfrak{A}$, $E \subset \mathfrak{M}$, where T is a constant.

Processes of such kind have been studied by Feller (1) and the author (2). In this general way of stating the problem I have followed A. N. Kolmogoroff (3).

Denote by $\pi(t, x, \tau)$ the probability that the state x of the system at moment t rests invariant in the time interval $t \leqslant s < \tau$. The function $\pi(t, x, \tau)$ satisfies, obviously, the identity

$$\pi(t, x, \tau) = \pi(t, x, \tau + \Delta\tau) + \pi(t, x, \tau)[1 - \pi(\tau, x, \tau + \Delta\tau)].$$

On the other hand, the following equality holds in virtue of (1):

$$1-\pi(\tau, x, \tau+\Delta\tau)=p(\tau, x)\Delta\tau+\Delta\tau\varepsilon+\delta$$

where

$$\delta \leqslant (2K\Delta\tau)^2+(2K\Delta\tau)^3+\ldots$$

Hence it is not difficult to verify that $\pi(t, x, \tau)$ is continuous with respect to τ and satisfies almost everywhere the equation

$$\frac{\partial\pi(t, x, \tau)}{\partial\tau}=-\pi(t, x, \tau)\,p(\tau, x)$$

Consequently, the function $\pi(t, x, \tau)$ is determined by the equality

$$\pi(t, x, \tau)=e^{-\int_t^\tau p(s, x)\,ds} \tag{2}$$

Put $\mathrm{E}(x, E)=1$ or $\mathrm{E}(x, E)=0$, according as $x \subset E$ or $x \subset \mathfrak{A}-E$. Now denote by $F_n(t, x, \tau, E)$ the probability that the system $\mathfrak{R}$, being in the state x at moment t, changes its state n times during the time from t to τ, so as to assume a state belonging to the set E at moment τ. The functions $F_n(t, x, \tau, E)$ are determined by the equalities

$$F_0(t, x, \tau, E)=\mathrm{E}(x, E)\,\pi(t, x, \tau)$$

$$F_n(t, x, \tau, E)=\int_t^\tau \pi(t, x, s)\,p(s, x)\,ds \int_{\mathfrak{A}_y} P(s, x, d\mathfrak{A}_y)\,F_{n-1}(s, y, \tau, E) \tag{3}$$

where the second integral is an abstract Lebesgue-Stieltjes integral defined with respect to the set family $\mathfrak{M}$ $(n=1, 2, \ldots)$. Analogous integrals will occur below. These formulae can be verified immediately.

Let a fixed set $e \subset \mathfrak{M}$ be given. Our purpose is to determine the probability $\Omega(t, x, \tau, E, \lambda)$ for the occurrence of the event A which consists in that the system $\mathfrak{R}$, being in the state x at moment t, alters in such a way that: 1) the state of $\mathfrak{R}$ at moment τ belongs to the set E; 2) the measure in the sense of Lebesgue of the time in the interval $t \leqslant s < \tau$, at which the state of $\mathfrak{R}$ belongs to e, exceeds λ.

Let further $\omega_n(t, x, \tau, E, \lambda)$ denote the probability that the event A will occur together with another event that consists in exactly n alterations of the state during the time $t \leqslant s < \tau$.

Then we put

$$\omega_n'(t, x, \tau, E, \lambda)=\mathrm{E}(x, e)\,\omega_n(t, x, \tau, E, \lambda)$$

$$\omega_n''(t, x, \tau, E, \lambda)=\mathrm{E}(x, \mathfrak{A}-e)\,\omega_n(t, x, \tau, E, \lambda) \quad (n=1,2,\ldots)$$

The following equalities give the solution of the problem stated above:

$$\omega_0(t, x, \tau, E, \lambda)=\mathrm{E}(x, eE)\,\pi(t, x, \tau)$$

$$\omega_n=\mathrm{E}(x, e)\,\omega_n'+\mathrm{E}(x, \mathfrak{A}-e)\,\omega_n'' \quad (n=1, 2, \ldots)$$

$$\omega_n'(t, x, \tau, E, \lambda)=$$

$$=\int_t^{t+\lambda} \pi(t, x, s)\,p(s, x)\,ds \int_{\mathfrak{A}_y} P(s, x, d\mathfrak{A}_y)\,\omega_{n-1}(s, y, \tau, E, \lambda-s+t)+$$

$$+\int_{t+\lambda}^{\tau} \pi(t, x, s)\,p(s, x)\,ds \int_{\mathfrak{A}_y} P(s, x, d\mathfrak{A}_y)\,F_{n-1}(s, y, \tau, E)$$

where $x \subset e$. $n=1, 2, \ldots$;

$$\omega_n'' (t, x, \tau E, \lambda) =$$

$$= \int_t^{\tau-\lambda} \pi (t, x, s)\, p (s, x)\, ds \int_{\mathfrak{A}_y} P (s, x, d\mathfrak{A}_y)\, \omega_{n-1} (s, y, \tau, E, \lambda)$$

where $x \subset \mathfrak{A} - e$, $n=1, 2, \ldots$;

$$\Omega (t, x, \tau, E, \lambda) = \sum_{n=0}^{\infty} \omega_n (t, x, \tau, E, \lambda) \qquad (4)$$

The validity of these formulae can be established by immediate verification. Using the induction method we can easily obtain the estimate:

$$0 \leqslant \omega_n (t, x, \tau, E, \lambda) \leqslant \frac{K^n (\tau - t)^n}{n!}$$

($n=0, 1, 2, \ldots$) whence follows that the series (4) converges absolutely and uniformly.

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Received
8.VI.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

MATHEMATICS

ON OPTIMAL COALESCENT TABLES OF SQUARES AND CUBES

By L. NEUSHULER

(Communicated by S. L. Soboleff, Member of the Academy, 26.IX.1944)

In our note [1] we considered the questions connected with the composition of optimal coalescent multiplication tables, *i. e.* such tables, in which the results are formed by means of «coalescent addition» (the writing one after another of separate groups of ciphers found in the table), for instance,2),3),4).Such coalescent tables may be composed also for powers, in particular, for squares and cubes, with a considerable economy in space and a sufficiently simple technique of operation. Moreover, if approximate results only are needed, the operation may be further simplified at the expense of accuracy.

The principal steps in the technique of operation with coalescent tables are: 1) the order, 2) the translation, and 3) the dimension, *i. e.* the number of lines, to which belong the elements (numbers) out of which the result is formed.

The proposed constructions for a given technique of operation possess the smallest possible volume. Also the distance (in the average) between the groups of ciphers out of which the result is formed I strived to make as small as possible.

All constructions of the tables of squares given here are based on the use of the following two properties of squares: 1) the first (senior) n ciphers of the number N determine up to two units in the last digit the senior $k=n+E\ 2\lg N-2E\lg N-1$ ciphers of its square, and 2) from $a\equiv b \pmod{5\cdot 10^{n-1}}$ follows $a^2\equiv b^2 \pmod{10^n}$.

For squares of five-digit numbers the optimal one-dimensional coalescent table of the second order with translation 0 is the table of the form (A) (here we have the scheme of the first page with the first two lines computed).

Four ciphers (two of which are placed on the top), marking the lines on the left (vary from 0001 to 5000) and on the right (from 9999 to 5000), are the junior four ciphers on N. The ciphers of the columns (the senior cipher of N) to the left of the rule in every column are to be composed with the groups of ciphers marking the lines on the left; and the cipher to the right of the rule is to be composed with the groups of ciphers marking the lines on the right. In the intersection of columns and lines the senior 4—5 ciphers of N^2 are located. The junior 5 ciphers of N^2 are located in the same line and the same column as the senior ciphers, yet their place is under the rule.

Examples: $10002^2=1000\ 40004$; $60002^2=36002\ 40004$; $39998^2=15998\ 40004$; $89998^2=80996\ 40004$.

The degree of tabularity of this table is $\cong 2$. It is possible to increase the degree of tabularity of the table (A) by taking out (in every line) 4 of the 5 junior ciphers in all columns under the rule, as well as a part of the senior ciphers. But this would increase the order and the dimension. An example of an optimal coales-

cent one-dimensional table of order 3 with translation 2 is presented by the author's table ([5]). Its degree of tabularity is $\cong 4$. As an example of an optimal coalescent two-dimensional table of order 3 with translation 2 may serve the table of complete squares of five-digit numbers, which consists of 1000 rectangles of the form (B).

Every rectangle corresponds to a group of five-digit numbers with fixed three middle digits (from 000 to 999), printed conspicuously above the first (from left to right) column. The remaining 9 columns are marked by ciphers (also printed in large characters) 1, 2, 3, . . ., 9 representing the cipher of units of N. The lines (of which there are 5) are denoted by pairs of ciphers representing the senior cipher of N. Under the first column are pairs of numbers (separated by a point) representing the first (1—4) ciphers of N^2. The numbers to the left (resp. right) of the point relate to five-digit numbers whose senior cipher is equal to the left (resp. right) cipher (of the pair), denoting the given line. Four-digit numbers in smaller print standing at the ciphers marking the remaining 9 columns, give the junior 4 ciphers of N^2. These numbers are printed half and half, one half above the other, and are to be read in the following order: the two-digit number printed below should be followed by the two-digit number printed above. The remaining (middle) 2 ciphers of the squares are chosen from the three-digit numbers printed at the intersection of the line marked by the senior cipher, and the column marked by the junior cipher, according to the following rule:

If the senior cipher of N marking the line is printed on the left, then from the corresponding three-digit number are to be chosen the first and second ciphers counted from left to right. If, on the other hand, the senior cipher is printed on the right, then the first and the second ciphers counted from right to left are to be chosen.

Remark 1. If the first of the two ciphers chosen from the table according to this rule is printed in bold type, then the number formed by the first group of ciphers is to be increased by 1. If the same cipher is printed in italics, then the number formed by the first group is to be increased by 2.

Remark 2. If N, instead of five valid ciphers, contains but four, three, two or one, then we write to it on the left correspondingly one, two, three or four zeros, and from the three middle digits find the corresponding rectangle.

Examples: $01234^2 = 1522756$; $51234^2 = 2624922756$; $41232^2 = (16_{99}+1)\,077824 = 1700077824$; $91231^2 = (38_{22}+1)\,095361 = 8323095361$; $91237^2 = (83_{22}+2)\,190169 = 8324190169$.

The degree of tabularity of this table is $\cong 5$.

It is possible to construct tables of squares with still higher degrees of tabularity: 1) by increasing the order, 2) by allowing a translation > 2, and by increasing both the order and the translation.

The possibility of further reducing the volume of these tables of squares through increasing the order, while leaving the value of the translation unchanged, is based on the fact that the group of ciphers determined by the senior $n-1$ ciphers of the n-digit numbers, and the group of ciphers determined by the junior $n-1$ ciphers of the n-digit numbers can be subdivided into subgroups composed by a smaller number of ciphers, and each of these subgroups can be tabulated separately.

Remark. When the groups of ciphers determined by the senior and junior ciphers of the n-digit numbers, for complete squares of which a table of a given order is built up, are subdivided into subgroups, the subdivision is to be chosen such as to make the degree of tabularity a maximum.

It may, however, be easily seen that no significant reduction of volume is obtained by thus increasing the order of the table.

A very effective means to reduce the volume of the table is to admit a translation $= 99$. The construction of a 2-dimensional table of order 3 with translation $= 99$

is examplified by the table of complete squares of five-digit numbers of the following form (C) (here we have the scheme of the first page with the first ten lines computed).

This table is arranged in the following manner: each of the 20 pages contains 20 columns. The left 9 columns are marked by the senior cipher of N. The right 10 columns are marked by the junior (fifth) cipher of N. The 3 middle (located in the column between the left and the right group of columns) ciphers of N (the senior of which is always placed on top) denote the lines.

The lines fall into groups of 10. In the seven-digit and eight-digit numbers situated in the left columns we distinguish three groups of ciphers: 1) the first 3 or 4 from left to right (of which the last two are printed in small type); 2) the following 2 printed in larger type; and 3) the last two printed again in small type.

The first two groups of ciphers are the senior ciphers of N^2, the third group is the translation. The four-digit numbers of the right 10 columns are the junior ciphers of N^2. The square of a five-digit number $a_1\, a_2\, a_3\, a_4\, a_5$ is determined in the following way: In the column (of the left group) a_1 in the line $a_2\, a_3\, a_4$ we take the number formed by the first two groups of ciphers. To it we add the number found in the same column a_1 in the line $a_2\, a_3\, a_4$, as formed by the third group of ciphers. The number thus obtained is prolonged by writing on the right-hand side of it the four-digit number found in the line $a_2\, a_3\, a_4$ at the intersection with the column a_5 on the right-hand side.

R e m a r k. If the number of the third group is printed in bold type, then the number of the second group is to be increased by 1.

In conclusion we describe the construction of the table of complete squares of six-digit numbers with the degree of tabularity $\cong 17$. This reduction of space (17 times) is achieved by increasing the order and the translation.

This table consists of: 1) 1000 rectangles of the form (D) (with fixed second, third, and fourth cipher of the six-digit number), and 2) a small tablet of the form (G), which for convenience is to be printed on a separate sheet.

The complete square of the six-digit number $a_1\, a_2\, a_3\, a_4\, a_5\, a_6$ is formed in the following way:

In the table (D): 1. From the line a_1 of the column $a_2\, a_3\, a_4$ of the lower tablet (a) write out either the number printed to the left of the point, or the number printed to the right of the point according as the position of the cipher a_1 in column (a) is to the left or to the right of the point. This is the first part of the result.

2. From the same line of the tablet (a) write out from the four-digit number of the column a_5 (large print) either the first, or the second, or the third cipher (if the cypher a_1 is printed in the column (a) to the left of the point), or, the fourth, or the second, or the third cipher (if the cipher a_1 is printed in the column (a) to the right of the point). This is the second part of the result.

3. From the upper tablet (b) from the line a_6 and the column a_1 write out the number formed by the two extreme ciphers printed in small type (they represent the translation), which we add to the coalescent sum of the numbers found in 1 and 2.

4. The so obtained number is coalescently summed with the two-digit number found in the same upper tablet (b) and in the same line (as in 3) a_6 in the column a_5. The number thus obtained represents the square of the number $a_1\, a_2\, a_3\, a_4\, a_5\, a_6$ without the three junior ciphers. These three junior ciphers are evidently determined by the last three ciphers of the number $(a_4\, a_5\, a_6)$, and are found in the tablet (G) whose construction and the use of which is as follows.

The ciphers (of which there are four arranged two and two one pair above the other), denoting the columns, represent the senior of the three ciphers, and the two-digit numbers, denoting the lines on the left (from 01 to 49) and on the right (from 99 to 50), represent the junior two ciphers. The senior ciphers printed below are to be combined with the junior ones printed on the left, and those printed above, with those printed on the right. The required last three ciphers of N^2 are situated at the intersection.

00	1 6	— 3 8	2 7	— 2 7	3 8	— 1 6	4 9	— 5	5	—
01	1_{000} 36_{001} 20_{001} 15_{999} 80_{998}	4_{000} 49_{001} 40_{001} 8_{999} 63_{998}	9_{000} 64_{001} 60_{001} 3_{999} 48_{998}	16_{000} 81_{001} 80_{001} 35_{998}	25_{001} 00_{001}	
02	1_{000} 36_{002} 40_{004} 15_{998} 80_{996}	4_{000} 49_{002} 80_{004} 9_{998} 63_{996}	9_{001} 64_{003} 20_{004} 3_{999} 48_{997}	16_{001} 81_{003} 60_{001} 35_{997}	25_{002} 09_{004}	
.						
.						
.						
60						

	1 2 3	1^{61}_{53}	2^{24}_{78}	3^{89}_{02}	4^{56}_{27}	5^{25}_{52}	6^{96}_{76}	7^{69}_{01}	8^{44}_{26}	9^{21}_{51}
0·5	$1{\cdot}26_{24}$	516	517	528	529	520	521	532	533	534
1·6	$126{\cdot}37_{49}$	132	152	184	206	227	248	279	290	312
2·7	$450{\cdot}50_{73}$	758	799	841	882	924	965	017	058	099
3·8	$975{\cdot}65_{98}$	374	436	508	569	621	682	754	716	877
4·9	$1699{\cdot}83_{22}$	990	072	164	246	328	400	491	573	655

(B)

	3·8 2·7 1·6 5 4·9 1·6 2·7 3·8 4·9 5	
01	201 401 601 801 001	99
02	404 804 204 604 004	98
.	. . . . .	.
.	. . . . .	.
.	. . . . .	.
50	500 500 500 500 500	50

1	2	3	4	5	6	7	8	9	0	0	1	2	3	4	5	6	7
1_{0000}	4_{0000}	9_{0000}	16_{0000}	25_{0000}	36_{0000}	49_{0000}	64_{0000}	81_{0000}	00	0000	0001	0004	0009	0016	0025	0036	004
$1_{0020}02$	$4_{0040}04$	$9_{0060}06$	$16_{0080}08$	$25_{0100}10$	$36_{0120}12$	$49_{0140}14$	$64_{0160}16$	$81_{0180}18$	01	0100	0121	0144	0169	0196	0225	0256	028
$1_{0040}04$	$4_{0080}08$	$9_{0120}12$	$16_{0160}16$	$25_{0200}20$	$36_{0240}24$	$49_{0280}28$	$64_{0320}32$	$81_{0360}36$	02	0400	0441	0484	0529	0576	0625	0676	072
$1_{0060}06$	$4_{0120}12$	$9_{0180}18$	$16_{0240}24$	$25_{0300}30$	$36_{0360}36$	$49_{0420}42$	$64_{0480}48$	$81_{0540}54$	03	0900	0961	1024	1089	1156	1225	1296	136
$1_{0080}08$	$4_{0160}16$	$9_{0240}24$	$16_{0320}32$	$25_{0400}40$	$36_{0480}48$	$49_{0560}56$	$64_{0640}64$	$81_{0720}72$	04	1600	1681	1764	1849	1936	2025	2116	220
$1_{0100}10$	$4_{0200}20$	$9_{0300}30$	$16_{0400}40$	$25_{0500}50$	$36_{0600}60$	$49_{0700}70$	$64_{0800}80$	$81_{0900}90$	05	2500	2601	2704	2809	2916	3025	3136	324
$1_{0120}12$	$4_{0240}24$	$9_{0360}36$	$16_{0480}48$	$25_{0600}60$	$36_{0720}72$	$49_{0840}84$	$64_{0960}96$	$81_{1080}08$	06	3600	3721	3844	3969	4096	4225	4356	448
$1_{0140}14$	$4_{0280}28$	$9_{0420}42$	$16_{0560}56$	$25_{0700}70$	$36_{0840}84$	$49_{0980}98$	$64_{1120}12$	$81_{1260}26$	07	4900	5041	5184	5329	5476	5625	5776	592
$1_{0160}16$	$4_{0320}32$	$9_{0480}48$	$16_{0640}64$	$25_{0800}80$	$36_{0960}96$	$49_{1120}12$	$64_{1280}28$	$81_{1440}44$	08	6400	6561	6724	6889	7056	7225	7396	756
$1_{0180}18$	$4_{0360}36$	$9_{0540}54$	$16_{0720}72$	$25_{0900}90$	$36_{1080}08$	$49_{1260}26$	$64_{1440}44$	$81_{1620}62$	09	8100	8281	8464	8649	8836	9025	9216	940

(b)

9	57_{20}	$04_{38}{\cdot}50_{56}$	97_{74}	44_{92}	92_{10}	39_{28}	87_{46}	36_{64}	
8	$33_{17}{\cdot}79_{33}{\cdot}26_{49}{\cdot}73_{65}{\cdot}20_{81}{\cdot}67_{97}{\cdot}14_{13}{\cdot}62_{29}{\cdot}10_{45}{\cdot}$								
7	$08_{15}{\cdot}54_{29}{\cdot}01_{43}{\cdot}48_{57}{\cdot}95_{71}$	$42_{85}{\cdot}90_{99}$	$37_{13}{\cdot}85_{27}$						
6	83_{13}	$30_{25}{\cdot}76_{37}$	23_{29}	70_{61}	$17_{73}{\cdot}65_{85}$	$12_{97}{\cdot}60_{09}$			
5	59_{11}	$05_{21}{\cdot}52_{31}$	99_{41}	46_{51}	93_{61}	40_{71}	88_{81}	36_{91}	
4	$34_{08}{\cdot}80_{16}{\cdot}27_{24}{\cdot}74_{32}{\cdot}21_{40}{\cdot}68_{48}{\cdot}15_{56}{\cdot}63_{64}{\cdot}11_{72}{\cdot}$								
3	$09_{06}{\cdot}56_{12}{\cdot}02_{18}{\cdot}49_{24}{\cdot}96_{30}$	$43_{36}{\cdot}91_{42}$	$38_{48}{\cdot}86_{54}$						
2	85_{04}	$31_{08}{\cdot}78_{12}$	$24_{16}{\cdot}71_{20}$	$19_{24}{\cdot}66_{28}$	$13_{32}{\cdot}61_{36}$				
1	60_{02}	$07_{04}{\cdot}53_{06}$	$00_{08}{\cdot}47_{10}$	94_{12}	41_{14}	89_{16}	36_{18}		

(a)

	1 2 3	1^{61}_{3}	2^{24}_{8}	3^{89}_{2}	4^{56}_{7}	5^{25}_{2}	6^{96}_{6}	7^{69}_{1}	8^{44}_{6}	9^{21}_{1}
0·5	$1{\cdot}26_{24}$	5156	5177	5208	5229	5250	5271	5302	5323	5354
1·6	$126{\cdot}37_{49}$	1352	1573	1804	2026	2257	2478	2709	2920	3152
2·7	$450{\cdot}50_{73}$	7558	7979	8401	8822	9254	9675	0107	0528	0959
3·8	$975{\cdot}65_{98}$	3754	4376	5008	5629	6251	6872	7504	8126	8757
4·9	$1699{\cdot}83_{22}$	9950	0772	1604	2426	3258	4070	4901	5723	6555

(D)

	99	1_{271}	2_{488}	3_{657}	4_{784}	5_{875}	6_{936}	7_{973}
0	—							
1	7	892_{485}	904_{383}	916_{293}	928_{215}	940_{149}	952_{095}	964_{053}
2	26	757_{728}	784_{575}	811_{440}	838_{323}	865_{224}	892_{143}	919_{080}
3	63	568_{971}	616_{767}	664_{587}	712_{431}	760_{299}	808_{191}	856_{107}
4	124	326_{214}	400_{959}	475_{734}	550_{539}	625_{374}	700_{239}	775_{134}
5	215	029_{457}	137_{151}	244_{881}	352_{647}	460_{449}	568_{287}	676_{161}
6	341	678_{700}	825_{343}	972_{028}	118_{755}	265_{524}	412_{335}	559_{188}
7	510	273_{943}	465_{535}	657_{175}	848_{863}	040_{599}	232_{383}	424_{215}
8	726	815_{186}	057_{727}	300_{322}	542_{971}	785_{674}	028_{431}	271_{242}
9	997	302_{429}	601_{919}	901_{469}	201_{079}	500_{749}	800_{479}	100_{269}

Remarks to the table (D). 1. If the number representing the translation (in the tablet (b)) is printed in bold type, then the first cipher of the second part of the result is to be increased by 1.

2. Remark 1 to construction (B) relates also to the tablet (a).

3. In determining the squares of five-digit numbers by the table (D) only the tablet (a) is to be used. It differs from the similar tablet of the construction (B) only in that the last (third) part of the result consists of three (and not of four) ciphers, and the fourth (from the end) cipher is included into the second part of the result (it is repeated within the limits of one column in all 5 lines). This is done for the convenience of determination of squares of six-digit numbers.

In every table of type (D) for certain values of the ciphers a_5 and a_6 the coalescent sum of the first two parts of N^2 with a corresponding translation is to be increased by 1 again. This does not depend upon a. In our example this circumstance is marked out in part (b) by points to the right of the numbers at the intersection of the corresponding lines and columns.

Since the properties similar to those which we used in constructing the tables of squares hold also for higher powers, it is possible to construct such tables for the latter also.

As an example of the possible constructions I will here give the coalescent two-dimensional table of cubes (of four-digit numbers) of the third order with the translation 3* (cf. (H)). It is to be used in the same way as construction (B). The only difference is that in it every line corresponds only to a single value of the senior cipher, and the second part of the result contains not 2 (as in the case of squares), but 6 ciphers, which decreases its degree of tabularity (the latter is here $\cong 2$).

Received
26.IX.1944.

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* The translation 3 occurs very seldom.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

HYDRODYNAMICS

A CONTRIBUTION TO THE STATISTICAL THEORY OF TURBULENCE

By V. G. NEVZGLJADOV

(Communicated by V. I. Smirnov, Member of the Academy, 25.VI.1944)

§ 1. O. Reynolds' equations (1) for the turbulent motion of an incompressible viscous fluid may be written, as has been pointed out in an earlier paper (2), in the following compact form:

$$\rho \frac{D\vec{\bar{c}}}{Dt} = \rho \vec{F} - \nabla(\bar{p} + \Pi) + k\Delta \vec{\bar{c}} - \frac{\partial \pi'_{ik}}{\partial x_k} \tag{1}$$

$$\operatorname{div} \vec{\bar{c}} = 0 \tag{2}$$

$$\frac{3}{2}\frac{D\Pi}{Dt} = k\Psi - \operatorname{div}\vec{P} - \Pi'_{ik}\frac{\overline{\partial u_k}}{\partial x_i} - \operatorname{div}\vec{I}. \tag{3}$$

Herein the following notations are adopted:

$$\Pi \equiv \frac{1}{3}\rho \overline{c'^2};\ \vec{I} \equiv \frac{\rho}{2}\overline{\vec{c}'c'^2};\ \Pi'_k \equiv \overline{\rho u'_i u'_k} - \Pi\delta_{ik};\ \vec{P} \equiv \overline{\vec{c}'p};\ \Psi \equiv \overline{(\vec{c}', \Delta\vec{c}')} \tag{4}$$

where $\vec{c} - \vec{\bar{c}} = \vec{c}'$ is the relative velocity, and u'_i, its components. Equations (1)—(3) are said to be fundamental equations. Their number is smaller than the number of functions involved. The principal problem of the theory of turbulence is to supplement the fundamental equations so as to make possible the determination of all quantities entering into them.

The author has solved this problem by the method of state equations, based on distinction between two groups of averaged quantities: fundamental quantities (the number of which is equal to the number of fundamental equations), as such were chosen $\bar{u}_i$, $\bar{p}$, Π, and the remaining quantities Π'_{ik}, $\vec{I}$, $\vec{P}$, Ψ called reducible. The reducible quantities can be expressed as functions of the fundamental ones, and these expressions are said to be dynamic state equations. The state equations were established in first approximation, whereby a complete system of equations was constructed on phenomenological considerations, supported afterwards by the kinetic equation of turbulent transfer.

The problem to be solved in the present paper is to supplement the fundamental equations with a statistical method.

§ 2. We assume the existence of an *a priori* density f_a for the probability of the mechanical state of the flow. We will consider the empirical probabilities for the observation time τ which we fix as a certain characteristics. Grouping the «almost equal» empirical probabilities, we form from them mean functions F, which possess a certain statistical weight W. The *a priori* probability f_a may be replaced with a satisfactory approximation by that function F which has the maximum weight. This leads to a variational problem whose object is to complete the fundamental equations by deducing the state equations from F.

Attempts to apply the variational principle to the construction of a statistical theory of turbulence, made by Karman ([3]) and Bürgers ([4]), gave no definite result, because the requirement imposed upon the variational problem has not been stated with sufficient clearness. The variational problem depends on the method used in describing the mechanical state of the flow. According to the choice of the object of statistical investigation, *i. e.* of the arguments of the function F, the results obtained vary not merely in the degree of the convenience of the mathematical form, but also in generality. We confine ourselves in the sequel to the consideration of an infinitely small domain Ω_0 about the point (x) of the flow. The value of velocity at the point (x) and the values of its derivatives with respect to the coordinates will be called natural variables, describing the mechanical state of the domain Ω_0. In natural variables the variational problem leads to a rather cumbersome mathematical apparatus. To obviate this difficulty, we subject our considerations to yet another restriction by taking for arguments of the function F the values of the relative velocities. To be determined is

$$F=F(u_i',\ u_{ik}';\ (x));\ u_{ik}'\equiv\frac{\partial u_i'}{\partial x_k};\quad \int F d\gamma=1;\quad d\gamma=\prod_{i,k=1}^{3} du_i'\,du_{ik}' \tag{5}$$

The function F is normed at every point (x), i. e. x, y, z are the parameters of the function F. Transition to relative velocities implies a restriction, since by means of our function F no average velocity $\vec{c}$ can be formed, and so it remains invaried.

We extend the invariability to all fundamental quantities to achieve a greater simplicity of the variational problem and of the whole mathematical apparatus of the theory, at the cost, however, of some indefiniteness in the final result. Let us formulate the variational problem as follows: To find the maximum of the functional W or, which is the same, of $\ln W$ under the conditions that

$$\int F d\gamma=1 \tag{6}$$

$$\int F u_i' d\gamma=0 \quad (i,k=1,2,3) \tag{7}$$

$$\int F u_{ik}' d\gamma=0 \tag{8}$$

$$\frac{1}{2}\int F c'^2 d\gamma=\frac{3}{2}\rho^{-1}\Pi \tag{9}$$

$$\frac{3}{2}\frac{D}{Dt}\frac{\Pi}{\rho}-\frac{3k}{2\rho}\Delta\Pi+\frac{k}{\rho}\int u_{ik}'u_{ik}'F\,d\gamma+U_{ik}\int F u_i'u_k'\,d\gamma+\operatorname{div}\int\frac{1}{2}c'^2F\,d\gamma=0 \tag{10}$$

$$U_{ik}\equiv\frac{1}{2}\left(\frac{\partial\bar{u}_i}{\partial x_k}+\frac{\partial\bar{u}_k}{\partial x_i}\right)$$

The conditions (7), (8) follow from the physical meaning of the arguments u_i', u_{ik}'; they are of kinematic character. The condition (9) is the result of invariability of the fundamental quantities. The condition (10) is the fundamental equation (3), *i. e.* the energy equation in which we omitted $\operatorname{div}\vec{P}$ as a small quantity. The conditions represented by the remaining fundamental equations add nothing essentially new to the formulated variational problem (the continuity equation does not affect it at all), but increase the indefiniteness of the result. The fact that the energy equation (3) plays an essential rôle in the variational problem, whereas the momentum equation (1) plays no such rôle, *i. e.* the fact that these fundamental equations so unsymmetrically affect the variational method of determination of the distribution function, appears to have a deep physical meaning. Another instance of the absence of symmetry is the following fundamental fact: there exists an inner — «latent» — energy, but there is no inner or «latent» momentum. The energy equation (3) takes into account the interaction of three energetic fields: the energy of the averaged motion, the mean energy of turbulent

pulsations, and the heat energy. We take the functional W in the form

$$\ln W \cong -\int F \ln F d\gamma + \text{const} \tag{11}$$

Solving the problem by Lagrange's method, we find

$$F = A \exp\left\{-\frac{1}{2} M c'^2 - (\vec{\Lambda}, \vec{c'}) - \mu_{ik} u'_{ik} - \rho^{-1} k \mu u'_{ik} u'_k - \mu U_{ik} u'_i u'_k + \right.$$
$$\left. + \frac{1}{2} c'^2 (\vec{c'}, \nabla\mu)\right\} \tag{12}$$

Here $\vec{\Lambda}$, μ_{ik}, M, μ are Lagrange's multiplicators (functions of the point (x)) of the respective conditions (7)—(10). The condition (8) is evidently satisfied for $\mu_{ik}=0$ (this is just the essential simplification of the apparatus achieved by transition to the relative velocities).

The solution (12), as well as the very formulation of the variational problem, holds only for not too great gradients of the averaged quantities. It may, therefore, be written in the form

$$F = A e^{-\frac{1}{2} M c^2 - \rho^{-1} k \mu u'_{ik} u'_{ik}} \left\{1 - (\vec{\Lambda}, \vec{c'}) - \mu U_{ik} u'_i u'_k + \frac{1}{2} c'^2 (\vec{c'}, \nabla\mu)\right\} \tag{13}$$

Using the equations (6)—(9), we obtain the state equations in the form

$$\Pi'_{ik} = -\rho^{-1} \mu \Pi^2 \left(\frac{\partial \bar{u}_i}{\partial x_k} + \frac{\partial \bar{u}_k}{\partial x_i}\right); \quad \vec{I} = \frac{5}{2} \rho^{-2} \Pi^2 \nabla\mu; \quad \Psi = \frac{3}{2\rho} \Delta\Pi - \frac{9}{2} \frac{\rho}{k} \mu^{-1} \tag{14}$$

The state equations (14) are not complete, since they contain no expression for μ. We may determine μ by taking one of the coefficients of the phenomenological theory. Let us take, for instance, the coefficient of turbulent viscosity from the state equation

$$\Pi'_{ik} = -\varepsilon_1 \Pi \left(\frac{\partial \bar{u}_i}{\partial x_k} + \frac{\partial \bar{u}_k}{\partial x_i}\right) \tag{15}$$

established in [2]. Here ε_1 is a constant. Comparing (15) and (14), we obtain:

$$\mu = \varepsilon_1 \rho \Pi^{-1}; \quad \vec{I} = -\frac{5}{2} \varepsilon_1 \rho^{-1} \Pi \nabla\Pi; \quad \Psi = \frac{3}{2\rho} \Delta\Pi - \frac{9}{2k\varepsilon_1} \Pi \tag{16}$$

Substituting (15) and (16) into (1)—(3), we obtain a complete system of equations for the determination of the five fundamental quantities $\bar{u}$, $\bar{p}$, Π, containing one constant ε_1, the numerical value of which must be found experimentally. This system coincides with that established in the phenomenological theory [2]. The latter, however, involves four constants ε_1, ε_2, g_1, g_2 (div $\vec{P}$ being neglected), which now are reduced to one ε_1, namely

$$\varepsilon_2 = \frac{5}{2} \varepsilon_1 \rho^{-1}; \quad g_1 = \frac{9}{2k\varepsilon_1}; \quad g_2 = 0 \tag{17}$$

The boundary conditions have been established above. The established equations describe the «one-scale turbulence». The term «one-scale» means that the interval τ of averaging in time, as well as $\bar{l}$ — the linear dimension of the cell for space averaging (if it be necessary for unstationary processes), are the same for all quantities. These intervals vary, in general, from point to point (they are functions of the point). This definition holds also for the phenomenological theory (where it was formerly [2] adopted with a superfluous restriction).

In conclusion I wish to thank V. Smirnov, Member of the Academy, for the discussion of different questions which arose in connexion with the present paper.

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Received
25.I.1944.

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Volume XLVII, № 7

GEOPHYSICS

ON THE PROPAGATION OF SOUND IN TURBULENT ATMOSPHERE

By V. KRASSILNIKOV

(Communicated by O. J. Schmidt, Member of the Academy, 3.V.1944)

The author's purpose in the present paper is to interpret on the basis of the statistical theory of turbulence, the results of some observations on phase fluctuations of a sound-wave propagating in a turbulent atmosphere (1).

The whole space in the region we are interested in between the loudspeaker and the microphone is imagined to be filled by the non-uniformities of the velocity field, which are superimposed upon one another. This gives us the picture of atmospheric turbulence.

Let us examine the case where we have a loudspeaker and two microphones at a certain distance (base) from each other. The x axis is defined to be directed from the loudspeaker to the centre of the base $B\,(B \perp x)$. If c be the velocity of the sound; ω, the circular frequency; v, the pulsation velocity of the wind; L, distance between the loudspeaker and the centre of the base, then the phase variation for the centre of the base may, when allowance is made for the influence of the non-uniformity of the wind along the propagation path $L-M$, be described as follows:

$$\varphi = \omega \int_0^L \frac{dx}{c+v} \cong \left(\frac{\omega}{c^2}\right) \int_0^L v\,dx + \text{const}$$

Here we neglect the square of the ratio v/c, so that only the velocity components of the wind directed along the propagation path of the sound are taken into account.

Supposing the angle θ between the x axis and the ray $LM_1\,(LM_2)$ to be small, we will have for the phase difference between the microphones

$$\psi = \varphi_{M_1} - \varphi_{M_2} = \frac{\omega}{c^2} \int_0^L (v_1 - v_2)\,dx \text{ and } \overline{\psi^2} = \left(\frac{\omega}{c^2}\right) \int_0^L \int_0^L \overline{\Delta v_1\,\Delta v_2}\,dx_1\,dx_2$$

To take account of the correlation of pulsations at different points of the flow let us make use of the «$^2/_3$» law, which has been obtained independently by A. N. Kolmogoroff (2) and A. M. Obukhov (3). According to this law, the square of the velocity fluctuation difference at points 1, 2 varies directly as the $^2/_3$ power of the distance between these points:

$$\overline{(v_1 - v_2)^2} = C^2 r_{1,2}^{2/3} \tag{1}$$

where C is the characteristic of turbulence having a dimension of $\text{cm}^{2/3}\,\text{sec.}^{-1}$, and $r_{1,2}$ is the distance between the points 1, 2*.

* The mean square of the longitudinal difference of the pulsation velocity [B_{dd} in the notation of A. N. Kolmogoroff (2)] is alone taken into account, whereas the correction for the mean square of the transverse difference (B_{nn}) is taken to be small.

As the turbulence is assumed to be isotropic, we have $\overline{v_1^2} = \overline{v_2^2}$ and $\overline{v^2} - \overline{v_1 v_2} = 1/2\, C^2 r_{1,2}^{2/3}$. Using this relation, we get for $\overline{\Delta v_1 \Delta v_2}$ the following expression:

$$\overline{\Delta v_1 \Delta v_2} = C^2 \{[(x_1 - x_2)^2 + \theta^2 (x_1 + x_2)^2]^{1/3} - (x_1 - x_2)^{2/3} (1 + \theta^2)^{1/3}\}$$

where x_1 and x_2 are coordinates of the points 1, 2 along the ray.

For ψ^2 we shall have:

$$\psi^2 = \left(\frac{\omega}{c^2}\right)^2 C^2 \int_0^L \int_0^L \{[(x_1 - x_2)^2 + \theta^2 (x_1 + x_2)^2]^{1/3} - (x_1 - x_2)^{2/3} (1 + \theta^2)^{1/3}\}\, dx_1\, dx_2$$

The evaluation of this integral will give us for the mean square fluctuation of the phase the following formula

$$\sigma\varphi = \sqrt{\overline{\psi^2}} = \text{const} \left(\frac{\omega}{c^2}\right) C L^{1/2} B^{5/6} \qquad (2)$$

where the order of magnitude of const is unity.

It should be noted that the above considerations are true for the case of geometric acoustics. With low frequencies, when the diffraction of sound at the non-uniformities enters into play, the relation to frequency can be different and will probably vanish more rapidly.

The experimentally established relation between the phase fluctuations and the velocity of the wind (1) in the frequency range 2000 — 5000 cycles is very near to direct proportionality. This can be explained on the basis of formula (2), if we suppose that the turbulence characteristics C varies directly as the velocity of the wind. From the point of view of the statistical theory of turbulence this relation appears to be quite natural and is in agreement with the results obtained by measuring the turbulence in the layer adjoining the Earth's surface (4, 5).

The dependence of $\sigma\varphi$ on ω and on the base B estimated as $B^{5/6}$ also agrees with the experiment.

The described theory of phase fluctuations can be applied for the computation of the mean square variation of the phase. But then we have to make the essential assumption that for a small interval of time Δt the variations of the turbulence field in a system of coordinates, moving at the average speed of the wind, are small in comparison with the local variations of turbulence produced by the translation of all the scheme at the mean velocity of the wind.

Let us now consider the case where we have a loudspeaker and a microphone placed at a distance L from each other, and the direction of the wind is supposed to be perpendicular to the line LM. In contradistinction to the case with two microphones considered above, the direction of the wind plays here a very important rôle.

Making use of the relation (1) we, after some computations, shall obtain the following formula for the mean square variation of the phase:

$$\sigma\Delta_t \varphi = \sqrt{\overline{\Delta_t \varphi^2}} = \sqrt{3L}\, C l^{5/6} \left(\frac{\omega}{c^2}\right) \qquad (3)$$

where $l = v\Delta t$.

If the quantities entering into this formula be given values obtained by experiment, one may determine the value of C. In the experiment we have: $\sigma\Delta_t\varphi = 18°$; $\Delta t = 0.1$ sec.; $L = 140$ m; $f = 800$ cycles; $\bar{v} = 6.5$ m/sec. So we get $C = 10.9$ cm$^{2/3}$ sec.$^{-1}$.

According to the data obtained experimentally by Gödicke (4), the value of C at a height of 1.15 m from the ground ($\bar{v} = 0.65$ m per sec.) is $C = 2.05$. Formula (3) gives $C = 10.9$; this value, as regards its order of magnitude, is comparable to the value obtained for C by several other

methods, if the difference in the velocities of the wind be taken into account. A very close coincidence could hardly have been expected in the present case in view of the difference in the height of the investigated layer and in the character of local conditions.

The results obtained may, in my opinion, be of use in revealing the source of errors in the operation of sound locators. Suppose the direction to the source of sound to form an angle α with the direction of the base. Then in an undisturbed flow the phase difference at the ends of the base will be

$$\varphi = \frac{2\pi B}{\Lambda} \cos \alpha$$

where Λ is the wave-length. If now the observed phase will, under the influence of turbulence, suffer a casual variation $\delta\alpha$, then

$$\delta\alpha = \frac{\Lambda}{2\pi B} \frac{1}{\sin \alpha} \delta\varphi$$

For large values of α

$$\delta\alpha = \frac{\Lambda}{2\pi B} \delta\varphi$$

A similar relation will take place for mean square errors

$$\sigma\alpha = \frac{\Lambda}{2\pi B} \sigma\varphi$$

or, if use is made of formula (2)

$$\sigma\alpha = \frac{\text{const}}{2\pi} CL^{1/2} B^{-1/6}$$

It follows from this formula that $\sigma\alpha$ does not depend upon the wave-length Λ, and only slightly depends upon the base. Using the data of the experiment, one can determine the constant in this formula. Let us take the case where

$B = 1$ m, $L = 20$ m, $\overline{v} = 2.7$ m/sec., $f = 4000$ cycles, $\Lambda = 8.5$ cm, $\sigma\varphi = 24°$.

For these data we obtain const ~ 2. If we take now into consideration that $\sigma\varphi$, and accordingly $\sigma\alpha \sim \overline{v}$, and that in the experiment considered $\overline{v} = 2.7$ m/sec. and $L = 20$ m, then the mean square angular error of the sound locator, due to turbulence, will be described by the formula

$$\sigma\alpha° \cong 0.3\, B^{-1/6} \left(\frac{L}{20}\right)^{1/2} \frac{\overline{v}}{2.7} \qquad (4)$$

where B, L, $\overline{v}$ are given in metres. If we take, for instance, $B = 1$ m, $v = 2.7$ m/sec., then for $L = 2$ km we have $\sigma\alpha \cong 3°$; so in the case under consideration the error in direction finding will be ~ 104 m.

For frequencies as low as $200 - 500$ cycles, which is the usual working range of sound locators, this formula must be refined, because with such frequencies it is not legitimate to confine ourself to conclusions derived from geometrical acoustics. Sound rangers are known to work badly under unfavourable meteorological conditions, especially under a strong wind. In the light of the considerations described here this may largely be accounted for by the action of the atmospheric turbulence on the mean velocity of the sound along the way of its propagation.

In concluding, I wish to express my heartiest thanks to Professor M. A. Leontovich for his constant interest in this work.

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Received
3.V.1945.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

CHEMISTRY

ON THE FORMATION OF β, β′-DICHLORODIETHYLETHER IN THE SYNTHESIS OF 2-CHLOROETHANOL

By E. ŠILOV

(Communicated by A. N. Frumkin, Member of the Academy, 30. VIII. 1944)

β, β′-dichlorodiethylether is found in the high boiling fraction of by-products obtained in 2-chloroethanol prepared by Gomberg's method [1].

In accordance with the modern theory of addition across the double bound [2] β, β′-dichlorodiethylether must be assumed to form by way of immediate reaction between ethylene, chlorine and 2-chloroethanol according to the scheme:

$$C_2H_4 + Cl_2 + H_2O \rightarrow Cl\begin{matrix}\swarrow CH_2 \rightleftarrows CH_2 \searrow \\ Cl \qquad\quad H\end{matrix} OCH_2CH_2Cl \rightarrow$$

$$\rightarrow ClCH_2CH_2OCH_2CH_2Cl + H^+ + Cl^- \qquad (1)$$

The kinetic equation of the reaction must be

$$\frac{dX}{dt} = k'E\,[C_2H_4]\,[Cl_2] \qquad (2)$$

where X is the concentration of β, β′-ether, and E is the concentration of 2-chloroethanol.

The value of k' cannot be found directly. Yet it is possible to establish the relative rate of reaction in comparison with the rate of formation of 2-chloroethanol.

Indeed, the rate of formation of 2-chloroethanol [3], if the formation of β, β′-ether is taken into account, must obey the equation (3)

$$\frac{dE}{dt} = k\,[C_2H_4]\,[Cl_2] - k'E\,[C_2H_4]\,[Cl_2] \qquad (3)$$

From the equations (2) and (3)

$$\frac{dX}{dE} = \frac{k'E}{k - k'E}$$

After integrating

$$X = E_0 - E + \frac{k}{k'}\ln\frac{\frac{k}{k'} - E_0}{\frac{k}{k'} - E}$$

From this equation k'/k can be determined, if the final concentration X of β, β′-ether and the initial and final concentrations of 2-chloroethanol E_0 and E are known.

In order to determine the value of k'/k the water solutions of 2-chloroethanol were saturated with ethylene and chlorine. In the obtained solution the values of X and E were found by analysis.

In comparison with the conditions on the basis of which equations (2) and (3) were established, our system is complicated by the presence of gaseous phase. In this case the over-all rate of the process is limited factually by the velocity of solution of ethylene. But as I have shown in an unpublished work, the distribution of chlorine and ethylene between the products of the reaction is not altered on the transition from homogeneous to heterogeneous system.

The results are collected in the table.

Run No.	E_0	E	$E-E_0$	X	k'/k
6	0.88	1.73	0.85	0.024	0.021
12	0.92	1.42	0.50	0.0146	0.025
9	0.86	2.62	0.76	0.035	0.020
7	1.93	2.49	0.56	0.024	0.019
11	2.56	2.93	0.37	0.023	0.021
8	2.86	3.58	0.72	0.082	(0.032) } an oil was formed
10	3.94	4.42	0.48	0.147	(0.056) } in the solution
					0.021

The results of the runs 8 and 10 have not been considered in computing the mean value of k'/k, since owing to prolonged saturation a liquid phase separated. Reaction between ethylene and chlorine in inaqueous phase naturally increases the ratio of β,β′-ether to 2-chloroethanol ([4]).

The fact that the value of k'/k is not prone to change within a wide range of concentrations of 2-chloroethanol confirms the correctness of our initial assumptions.

2-chloroethanol is known to form according to a scheme analogous to equation(1), but the place of 2-chloroethanol is occupied by the H_2O molecule ([3]). Therefore the value of k'/k enables to compare the activity of molecules of H_2O and ClH_2CH_2OH in the reaction with C_2H_4 and Cl_2. It is to be taken in consideration that 1 l of 1 M solution of 2-chloroethanol contains about 54 mols water. This ratio of concentrations of water and 2-chloroethanol is very close to the ratio of velocities $k/k' = 48$. Thus, the molecules of water and 2-chloroethanol can roughly be said to be equally active in the two analogous reactions.

Experimental (the experiments were carried out in collaboration with E.A. Chistova). A tube 1 m high and 18 mm in diameter was used as a reaction vessel. The tube was provided below with two inlets for gases, placed one against the other, and with a stop-cock to let the solution out after the reaction. A side-tube 8 mm in diameter was soldered parallel to the reaction tube; its lower end was placed immediately underneath gas inlets, and its upper end was several centimetres below the upper edge of the reaction tube. Owing to the side-tube the liquid was set in continuous circular motion by the passage of gases. It had been proved by special experiments that the mixing attained in this way is quite satisfactory. In order to prevent photochemical chlorination the apparatus was completely protected by black photographic paper. The experiments were made at room temperature.

Ethylene and chlorine passed into the reaction tube with equal speed—20 ml/min., controlled by rheometers. 300 ml of solution were poured in the tube so that upon letting the gases in, the liquid rose to the top of the tube. In this way was prevented the possibility of the solution being contaminated by the products of the reaction of gases on the liquid films or on the wall.

In the interval during which the experiment lasted the absorption of ethylene and chlorine increased the volume of the solution by as much as 10—20 ml. This

surplus came out through a drawn tip of the tube into a glass attached to the side of the tube.

The saturation of the solution with gases continued 6—12 hours, during which time 50 to 70 per cent of the passed gases were absorbed. In blank experiments it was found that unabsorbed gases carry off relatively large amounts of dichloroethane, but neither β,β′-ether nor 2-chloroethanol are lost on account of their low vapour density ([5],[6]).

After the reaction the solution and washings were poured into a 500-ml measuring flask, and the latter was filled with water to the mark. β,β′-dichlorodiethylether, hydrochloric acid and 2-chloroethanol were determined in the solution.

For determination of β, β′-dichlorodiethylether a distilling apparatus was used built after a model of the well-known Dina-Stark device. β,β′-ether was distilled together with dichloroethane from 50—100 ml solution in a 1-ml burette. To remove some admixed 2-chloroethanol the distilled oil was returned in the bulb of distilling apparatus and redistilled from 100 ml pure water. The oil thus obtained was after measuring its volume treated with concentrated sulphuric acid in a pipette with two stop-cocks. β,β′-ether dissolved, and the remaining dichloroethane was measured in a burette. From difference the volume of β,β′-ether was found. The check analyses of known mixtures of β,β′-ether with 2-chloroethanol and dichloroethane have shown that the result must be corrected by addition of 0.05 ml With this correction the accuracy of determination of β,β′-ether is about ±0.05 ml. corresponding to a relative error of determination of X and k'/k equal 10—20 per cent.

Hydrochloric acid was determined by alkali titration using methyl-orange as indicator. 2-chloroethanol was determined by alkaline hydrolysis and argentometric titration. As dichloroethane is slowly hydrolysed by alkali, the hydrolysis has to be carried under strictly determined conditions.

We carried out the hydrolysis in 0.4 *M* natron for 25 minutes at 50° C.

The results of experiments are given in the table.

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Received
30.VIII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

PHYSICAL CHEMISTRY

MOLECULAR MECHANISM OF THE SOLIDIFICATION OF POLYMERS*

By S. N. ZHURKOV

(Communicated by A.F.Joffé, Member of the Academy, 29.VII.1944)

Very little has been done to elucidate the process by which high-molecular polymers pass over from a rubber-like to the solid state. The solidification of natural rubber (1) and polychloroprene on long storage is ascribed to the formation of a crystal phase within the polymer. However, such a mechanism is not all-embracing. As a matter of fact, most polymers (polystyrol, polybutadiene, polymethylmetacrylate, many co-polymers, and others) do not undergo phase transformations when solidifying, and the process in these cases is similar to the formation of an amorphous glassy mass by supercooled liquids (2).

Solidification of non-crystallizing polymers takes place in a comparatively wide range of temperatures and may be followed by direct measurements of their rigidity or determined from the discontinuity in the dependence of the specific volume upon the temperature. The temperature range in which one or another polymer solidifies is determined by the properties of the monomer molecules and by the shape and flexibility of the polymer chains they form, and also depends upon the quantity of low-molecular substances (plastifiers) dissolved in it. Furthermore, thanks to the relaxation mechanism of the change in properties in the solidification interval (3,4), the temperature of solidification is also determined by the duration of exposure to external influences.

In the present work an attempt is made to outline the molecular mechanism of the solidification of polymers on the basis of the results of an investigation of the phenomenon by the volume sorption of volatile solvent vapours (5). In this method the molecules absorbed by the polymer fulfil the function of test particles, enabling a determination of the part played by the different molecular groups of the polymer molecule in the solidification process. The fundamental conclusion arrived at as a result of the experiments, namely, that the fall in the temperature of solidification is proportional to the number of sorbed molecules irrespective of their chemical nature, points to the direct dependence of the process upon the concentration in the polymer of certain active groups which are blocked during sorption.

It is our opinion that the formation of molecular bonds between such active sites of the polymer molecules lies at the basis of the mechanism of the transition of polymers from a rubber-like to the solid state. These bonds, being sufficiently stable to rupture by molecular heat motion, play an especially important rôle in complex molecular interaction. They fix the positions of the polymer molecules with respect to one another, thereby forming nodal points in the macromolecular space-lattice.

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* Paper read at the second conference of high-molecular compounds convened by the Academy of Sciences of the USSR, May 22—25, 1944.

Owing to steric hindrances, such nodes form comparatively rarely. Their energy of formation, however, is not great and is comparable with the energy of heat motion. A change in temperature therefore changes the equilibrium concentration of the nodes. Heating of the polymer above a fixed temperature partially destroys the nodes linking the chain-line molecules and transfers the polymer into a rubber-like state. Cooling, on the other hand, leads to the formation of a sufficiently stable space-lattice, which lends the polymer the properties of a solid.

The natural conclusion is that the molecular bonds that form the nodes in the solidification range are subject to constant destruction by heat motion and rebuilding.

From this point of view the rigidity of a polymer should depend not only upon the node concentration, *i. e.*, upon the density of the lattice structure, but also upon how long each node continues in a bound-up state as compared to the interval during which the deforming force is applied (relaxation mechanism).

Applying the law of mass action to the node formation reaction and making use of experimental data on the lowering of the temperature of solidification during the sorption of vapours, we may compute the binding energy of a node*.

Let N be the concentration of active groups in a polymer, and x the number of nodes they form. Let us assume, further, that two groups are involved in the formation of each node. We may then write the following equation for such a bimolecular reaction:

$$x = (N - 2x)^2 Ae^{u/RT} \qquad (1)$$

where A is the steric factor, and u is the binding energy of a node.

An analogous equation may be written for the case where a polymer has adsorbed n molecules of a solvent, resulting in a fall in its solidification temperature to a value T_1. In conformance with our experimental results we make take it for granted that each solvent molecule adsorbed means the exclusion of one active group. Then we have:

$$x_1 = (N - 2x_1 - n)^2 Ae^{u/RT_1} \qquad (2)$$

We assume that the rigidity of a polymer, which is measured by its modulus of elasticity, is determined by the number of nodes x. Hence two states having identical values for the modulus of elasticity should have the same number of nodes.

Equating the right-hand sides of equations (1) and (2) and cancelling the steric factor, we get:

$$(N - 2x)^2 e^{u/RT} = (N - 2x - n)^2 e^{u/RT_1}$$

Here T_1 is the temperature of the polymer in the swollen state at which its modulus of elasticity is equal to that of the polymer in the pure state at temperature T.

Upon transposing, this expression assumes the following form:

$$\left(1 - \frac{n}{N - 2x}\right)^2 = e^{\frac{u}{R}\left(\frac{1}{T} - \frac{1}{T_1}\right)}$$

or

$$2 \log \left(1 - \frac{n}{N - 2x}\right) = \frac{u}{R} \frac{T_1 - T}{TT_1}$$

For low concentrations $n/N \ll 1$, and hence also $2x \ll N$. Therefore we may write N for $N - 2x$, and discard everything but the first term in the expansion of the expression $\log \left(1 - \frac{n}{N}\right)$. Since for small value of n/N, $T - T_1 = \Delta T$ is also small, we may write T^2 instead of TT_1. Thus the expression

* This computation was carried out in collaboration with E. V. Kuvshinsky.

for the binding energy of a node in a space lattice reduces to the following simplified form:

$$u = \frac{2nRT^2}{\Delta T N}$$

where T is the temperature when solidification sets in, corresponding to values of x that are close to zero.

The admissibility of the assumption that have been made ensues directly from experimentally established facts [5].

1) In the case of sorption the $\lg E = f(t)$ curves which describe the dependence of x upon t undergo a parallel displacement. These curves show that ΔT is independent of x and that its value may be extrapolated to the values of x close to zero.

2) The proportionality of ΔT to n justifies the replacement of TT_1 by T^2. This shows that $\Delta T N/n$ has the same value for all values of ΔT right down to the very smallest, when $T = T_1$.

The quantities T and $\Delta T/n$ have been experimentally determined for four synthetic polymers: polymethyl-metacrylate, a combination (co-polymer) of divinyl and the nitrile of acrylic acid, containing 33 per cent of the latter, divinyl rubber and polyisobutylene.

Upon making certain assumptions as to the relative nature of the molecular groups of the polymer that are involved in the formation of the lattice, we may compute their number N, and hence determine for each polymer the quantity u, *i. e.*, the binding energy of a node. The results of a calculation of the binding energy u are given in the accompanying table, together with a comparison with experimental values for the total binding energy U obtained from the heats of vaporization.

Polymer	Active group interacting in solidification	T°K	$\frac{\Delta T}{n}$	u calculated (in cal./mol group)	U exper. (from heat of vaporization)	U/u
Polymethyl-metacrylate	$-C(=O)-OCH_3$	360	230	2250	5600	2.5
Nitrile rubber	—CN	260	195	2220	—	—
Polyisobutylene	$-CH_3$	250	106	670	1780	2.7
Divinyl rubber	CH_2- ; $=CH-$	240	97	320	990	3.1

The fact that u is of the same order of magnitude as the energy of heat motion at the solidification temperature, and that the ratios of the values of u to the total values of U and the solidification temperatures T are regular, fits in well with the hypothesis on the mechanism of solidification that has been put forward. That the values of U are 2.5—3 times larger than the values of u is explained by the fact that U stand for the total binding energy during evaporation, whereas the values of u we have computed represent only that part of the energy U expended to rupture the bond between the groups and does not include the energy used up to remove them from the condensed system. The ratio of U to u is similar to that between the heat of vaporization and the heat of fusion.

In conclusion I consider it my pleasant duty to express my gratitude to A. F. Joffé for the interest he has shown in this work and for his valuable advice in carrying it out.

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Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

PHYSICAL CHEMISTRY

ON THE POISONING OF CATALYSTS

By S. Z. ROGINSKY, Corresponding Member of the Academy

That surface non-uniformity plays an important rôle in the poisoning of contacts had been known for more than 20 years, yet not until recently was a comprehensive theory of poisoning built up on this basis. This is easily done for a highly non-uniform surface by employing the leading (control) band method (1). The notion of contact poisoning is generally understood to embrace all cases where the decrease in catalytic activity is due to sorption of foreign matter. The mechanism of this phenomenon vary widely. In fact, the activity of the catalyst can diminish: 1) as a result of unfavourable changes in the adsorption and kinetic surface constant caused by the penetration of the poison into the surface layer of the catalyst lattice, or 2) the reason may lie in the poison being adsorbed by some isolated microscopic portions of the surface, which are thereby prevented to influence the reaction.

Such poisoning by microblocking of the surface is essentially different from 3) macroscopic blocking which is due either to the filling of the pores and capillaries by easily condensing liquids, or to the formation of a polymolecular barrier from the solid reaction products, which hampers access to the active surface.

The classical theory of the poisoning of catalysts with a uniform surface put forward by Langmuir is applicable only to the case of microblocking on uniform surfaces.

Below is developed a theory of microblocking for highly non-uniform surfaces (see also (2)).

Catalyst poisoning at constant temperature is conveniently characterized by the dimensionless quantity λ, which is the ratio of the activity of the poisoned catalyst A_Γ to its activity in the absence of contact poisons A_0:

$$\lambda = A_\Gamma / A_0 \tag{1}$$

λ is directly dependent upon the amount Γ of poison adsorbed by the contact or upon a proportional quantity Φ_Γ which is the portion of the total surface area occupied by the poison:

$$\Phi_\Gamma = \Gamma / \Gamma_\infty \tag{1a}$$

where Γ is the amount of adsorbed poison, and Γ_∞ is the value of Γ when the entire surface is covered by one layer of the poison. The function describing the dependence of λ upon Φ_Γ

$$\lambda = \lambda(\Phi_\Gamma) \tag{2}$$

may be called isotherm of poisoning with reference to the ratio of the poisoned area to total area or, simply, Φ_Γ [Γ]-isotherm. This function is applicable to all types of poisoning, reversible as well as irreversible. In the latter case use is frequently made of the dependence of the activity upon the concentration, or upon the partial pressure of the contact poison in the volume.

Of such concentration isotherms, the following will be used hereafter:

$$\lambda = \lambda(\pi) \tag{3}$$

where π is the partial pressure of the contact poison in the gas volume.

A. Linear Isotherms. $\Phi_\Gamma[\Gamma]$-isotherms and concentration isotherms are often found to have a linear course (see[3] and the earlier researches by Maxted). Usually they are regarded as a characteristic indication of surface uniformity, since, according to Langmuir's theory, only one equation is possible for these isotherms:

$$\lambda = \Psi_\Gamma \tag{4}$$

where Ψ_Γ is the poison free portion of the total surface area

$$\Psi_\Gamma = 1 - \Phi_\Gamma \tag{4a}$$

But in actual fact the surface must not necessarily be uniform in this case. Equation (4) will hold, indeed, for any degree of non-uniformity of the surface (Fig. 1) and for any position of the control band, provided that the arrangement of the molecules of the poison over the surface is in no way connected with its activity. There will be no such connexion, for instance, if the Q of the adsorption of the poison is not correlated with the E of the catalytic reaction, or again if there is no correlation between the E of the poison and the E of the catalytic reaction.

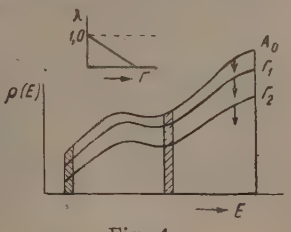

Fig. 1.

In the first case the final picture is represented in Fig. 1; in the second case a redistribution of the poison over the surface is possible with the result that the activity of the latter may either be reduced considerably (Fig. 2, *a*) or not reduced at all (Fig. 2, *b*).

Taking into account the possibility of a linear dependence between the residual activity and the quantity of adsorbed poison, owing to the adsorption of successive layers of the poison by porous and powder contacts, as was shown by Russell and Ghering (4), and the possibility of replacing most of the complex functions by linear ones when the interval in which the argument varies is limited, one has to admit that the linearity of $\Phi_\Gamma[\Gamma]$-isotherms is of itself no evidence of the uniformity or non-uniformity of the surface.

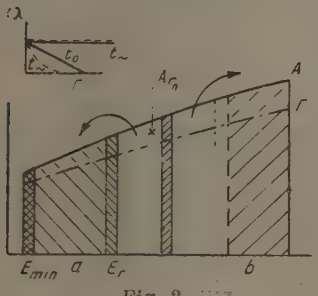

Fig. 2

B. Poisoned Area Isotherms when the Most Active Spots are Mainly Affected by the Poison. The situation is entirely different in the case of a preferential poisoning of the more active sites. This may occur because the value of the heat of adsorption is higher at these spots (E and Q vary in the opposite sense), or the adsorption of the poison is more rapid (both E's vary in the same sense), or, finally, because of the fixation of the poison at the spot where it is formed during auto-poisoning. In all these cases a simple relation should exist between the quantity of adsorbed poison and the change in the energy of activation of the catalytic process

$$\Phi_\Gamma \simeq \int_{E_{\min}}^{E_\Gamma} \rho(E)\, dE \tag{5}$$

with the equation

$$\rho(E) \simeq \rho(E_\Gamma) = d\Phi_\Gamma / dE_\Gamma \tag{6}$$

that follows from it (Fig. 3).

To derive the $\Phi_\Gamma[\Gamma]$-isotherm, we make use of the equation for the fractional activity λ' (5):

$$\lambda = \frac{A_\Gamma}{A_0} = \frac{M_\Gamma \exp(-E_\Gamma/R\Gamma)}{M_1 \exp(-E_1/R\Gamma)} \approx e^{-\gamma E_\Gamma/RT} \tag{1b}$$

By making use of equation (5) it is not difficult to express E_Γ through E_1 and Φ_Γ, which immediately gives the result sought for:

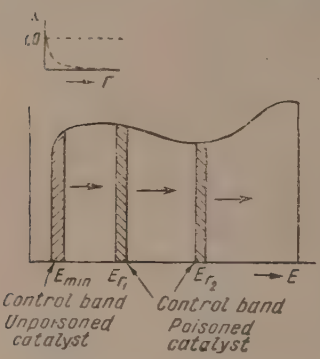

Fig. 3.

a) for exponential distribution:

$$E_\Gamma \simeq \frac{1}{\alpha} \ln \frac{\alpha \Phi_\Gamma}{H} \text{ when } \alpha > 0$$

b) for power distribution:

$$E_\Gamma \simeq \beta \Phi_\Gamma^{1/(n+1)}, \quad \beta = \sqrt[n+1]{\frac{n+1}{H}}$$

c) for uniform distribution:

$$E_\Gamma = E_1 + \frac{\Phi_\Gamma}{H} \tag{7}$$

This leads to the following equations for λ (in the case of exponential distribution allowance is made for the change of M_Γ with increasing E):

a) for exponential distribution:

$$\lambda \simeq \lambda_0 / \Phi_\Gamma^m, \quad m = \frac{1 - \alpha RT}{\alpha RT}$$

b) for power distribution:

$$\lambda \simeq \lambda_0' \exp[-\alpha \Phi_\Gamma^{1/(n+1)}], \quad \alpha = \beta / RT$$

c) for uniform distribution:

$$\lambda = \lambda_0'' \exp[-\alpha \Phi_\Gamma], \quad \alpha = 1/HRT \tag{8}$$

Of these equations c) was discovered by us (6) during a study of the catalytic oxidation of carbon dioxide (CO_2 *in statu nascendi* is the contact poison); equation a) seems to have been obtained by Kubokawa (7). Equation b) appears for the first time. These equations, as well as equation (6), make it possible to determine the form of the function $\rho(E)$. Taking (1b) into account, we may write equation (6) in a somewhat different form:

$$\gamma \rho_s^0(E) \simeq -\frac{1}{RTd} \frac{d\Phi_\Gamma}{\ln \lambda} \tag{9}$$

It may be remarked that when the poisoning is very slight, all of the isotherms conform to the linear equation $\lambda = \lambda_0(1 - c\Gamma)$. It is only when the degree of poisoning attains medium or high values that the differences between the isotherms, accounted for by differences in the type of the function $\rho(E)$, become significant. To study the nature of non-uniformity from the poisoning the interval of variation of λ must be sufficiently wide, in any case not narrower than two orders.

The equations that have been derived are applicable to many systems with a stationary control band lying at the distribution boundary. Here poisoning fails to alter the kinetics of the process. The same van't Hoff equations are retained in one limiting case, and the zero orders in the other. For systems with a mobile (suspended) band there is little probability that the blocking will start from the sites controlling the process. In this case poisoning should be better described by $\Phi_\Gamma[\Gamma]$-isotherms obtained for a poisoning beginning from slightly active sites.

C. Poisoning Beginning with Active Sites. In this case for systems in which the control band is fixed at the edge blocking by the contact, poison begins to involve all but the controlling sites which are the last to be blocked (Fig. 4). Therefore up to a certain value of Φ_r close to unity the surface shows no sign of poisoning at all, and not until the limits of $\Delta\Phi_r$ are such as to change E by as much as $(1\div2)\ RT$, does a complete poisoning take place (Fig. 4). In other words, a critical stage in the process of surface poisoning should exist, at which the activity changes abruptly from A_0 to 0.

In the case of a mobile band the picture is more complicated. If covering by the poison proceeds from small values of E, then up to a certain value of Φ_r we have:

$$A=\text{const}=A_0 \qquad (10)$$

and, from this value of Φ_r onwards, a rapid decline takes place. However, inasmuch as the band is not pressed to the edge, it begins to move towards large values of E, and poisoning obeys the law characteristic for the blocking of the most active sites, with different curves for different ρ. When the band reaches the edge, the activity falls rapidly to zero. Characteristic is the sharp change in the kinetics (Fig. 5) upon transition through Φ_r critical.

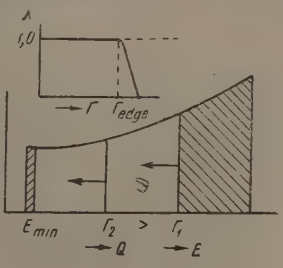

Fig. 4

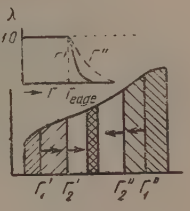

Fig. 5

Within definite limits the change in the concentration ceases to affect the rate. The fractional orders described in our preceding paper ([5]) are replaced by zero orders. However, upon a sufficiently marked lowering of the partial pressures, such as tears the control band away from the blocked zone, the fractional orders should again be re-established.

The picture of the poisoning process is different when the sites with maximum values of E are the first to be covered by the poison. The instant the poison reaches the leading band, the latter begins to move. At the same time the fractional orders are transformed not into zero orders, but into simple van't Hoff orders.

The true energy of activation does not increase with poisoning but rather declines. But since to the left of the leading band covering of the surface by the exothermic complex also decreases exponentially, and for Q faster than for E, the temperature coefficient at a given Φ_r is determined not by E but by $E-Q$, and the apparent energy of observation $E_{obs}=E-Q$ increases as the poisoning goes on.

Thus, during poisoning the equality $E_r=E$ is possible; here E is the energy of activation at the given sites, and also $E_r=E-Q$. The activation energies determined from the temperature coefficient during constant fractional covering always increase with the poisonig.

Such are the basic peculiarities of the isotherms of poisoning with respect to covering on broad non-uniform surfaces. Their study in a large interval of activities may give highly valuable information both on the statistics of the catalytic process and on the mechanism of the latter.

Received
20.III.1945.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

PHYSICAL CHEMISTRY

ON A THEORY OF THE PHENOMENON OF SPIN IN DETONATION

By K. I. SHTSHOLKIN

(Communicated by N. N. Semenov, Member of the Academy, 2. IV. 1945)

In certain cases the combustion zone in detonation front does not occupy the entire cross-section of the tube, but is concentrated in the form of a head near its periphery. Spreading forward, the head at the same time rotates around the tube axis with such a speed that the path of the flame forms an angle of about 45° with the cross-section of the tube. The pitch of the helix is proportional to the diameter of the tube (1).

The phenomenon described above has been called the spin in detonation, and is usually observed in combustible mixtures approaching in composition the limits of detonation, or in so-called slow-burning mixtures ($2CO + O_2$, C_5H_{12} + air, *etc.*). In all other mixtures if spin occurs at all, it is observed only near the spot where detonation originates (2). In several cases the spin structure of a fast flame has been detected in the pre-detonation period (2).

Notwithstanding the fact that the phenomenon of spin in detonation has been known for 20 years, it has not been explained.

The possible mechanism of spin we have proposed reduces essentially to the following: in spin detonation it becomes impossible for the gas mixture to ignite in the plane shock-wave (as is the case in ordinary detonation) owing to its comparative chemical inertness, and the mixture ignites thanks to strong perturbation of a special nature which spreads over the front of the shock-wave along the periphery of the tube.

This perturbation is a break in the front of the shock-wave which, as is not difficult to see, has a higher temperature and density than the plane shock-wave (the break in the wave is a case of the partial collision of two shock-waves). The break appears as a result of the origination of a circumferential component of the velocity of the products of combustion beyond the section of the flame oblique to the front of the shock-wave. The circumferential velocity component of the products, encountering the unburned gas moving behind the front of the shock-wave along the tube (near the front of the shock-wave the flame occupies only a part of the cross-section of the tube), changes the direction of its velocity, distorting in this way the sector of the shock-wave adjacent to the flame, and forms a corner in which the following portion of gas ignites. The break moves along the circumference with exactly the same velocity that the combustion products acquire over the circumference. The slope of the flame toward the front of the plane shock-wave develops as a result of the fact that the flame, appearing in the break of the shock-wave, lags behind the latter just as soon as the break (and ignition) shifts further along the circumference.

Now the circumferential component of the velocity of the combustion products causes the sector of the shock-wave to swerve —to break its front, and the break ignition to shift along the circumference; the shift in ignition along the circumference causes the flame to slope toward the front of the plane shock-wave and, finally,

the slope of the flame swerves the combustion products in the same direction giving rise to a circumferential component of the velocity, and thus the cycle is completed. The energy of perturbation is derived from the combustion products. In the detonation wave some kind of redistribution of energy takes place. In the process of combustion a vortex is formed. It may be assumed that by the time combustion is over, the vortex vanishes and that therefore the usual Jouguet equations will be applicable to the spin detonation if the conservation equations are written without taking into account the intermediate stages of combustion. By the way combustion of the gas beyond the head of the spin evidently takes place normally (heat conduction, mixing) along a rather extended surface.

If the gas flow intersects the flame at an angle α, the combustion products change direction and form an angle β with the flame.

Denoting the density of the gas before and after combustion by ρ_1 and ρ_2 respectively, we may write:

$$\operatorname{tg}\alpha/\operatorname{tg}\beta=\rho_2/\rho_1 \qquad (1)$$

This follows directly from the principle of continuity.

If we assume that the rate of shift in perturbation along the circumference of the tube is equal to the circumferential component of the velocity of the combustion products beyond the flame front which forms an angle α with the generatrix of the tube, it is not difficult, upon taking (1) into account, to find this angle:

$$\operatorname{tg}\alpha=\sqrt{\frac{\rho_1}{\rho_2}-2} \qquad (2)$$

where ρ_1 is the density of the gas in the shock-wave before combustion, and ρ_2, its density after combustion.

After this it is a simple matter to obtain the pitch of the helical path of the flame, p:

$$p=\frac{\pi d}{\operatorname{tg}\alpha}=\frac{\pi d}{\sqrt{\frac{\rho_1}{\rho_2}-2}} \qquad (3)$$

and the frequency per second of the spin, n:

$$n=\frac{D}{p}=\frac{D}{\pi d}\sqrt{\frac{\rho_1}{\rho_2}-2} \qquad (4)$$

where d is the diameter of the tube, and D is the speed of detonation.

Expressions (3) and (4) state that the pitch of spin is directly proportional, and the frequency per second of spin inversely proportional to the diameter of the tube, which is in agreement with experiment.

If ρ_1 and ρ_2 are computed in accordance with the classical theory, we get the following data, for example, for the mixture $2CO+O_2$ ($D=1795$ m/sec.) in a tube 2.55 cm in diameter:

$$\operatorname{tg}\alpha=\sqrt{\frac{5.13}{1.83}-2}\cong 0.90 \quad (\alpha\cong 42^\circ)$$

$$p=\frac{2.55}{0.90}\cong 8.9\,\text{cm}; \quad n=\frac{1795}{8.9\cdot 10^{-2}}\cong 20\,000$$

instead of $p=7.5$ cm and $n=24\,000$ found experimentally by Bone and Frazer [2] under the same conditions.

Computations of the characteristics of the spin for other mixtures with formulae (2)—(4) give similar agreement with experimental data, except when there is

a suggestion that several heads of spin are formed (the experimental data for methane are approximately an integral number of times (2—4) higher than the computed).

The spin mechanism makes it possible to explain experiments of a qualitative nature as well. We make special note of the fact that the perturbation (break) in the front of the plane shock-wave in the case of spin detonation was experimentally detected by Bone, Frazer and Wheeler ([3]), with the aid of instantaneous Toepler photography.

Finally, mention should be made of the fact that under the conditions given above (the mixture is ignited by the break in the wave, but not by the plane front) ignition at any spot in the front of the shock-wave is unstable: a slight shift in combustion to one or another side along the circumference sets the perturbation (ignition) in motion along the circumference. This makes it clear why the spin must originate when there is chance perturbation of the front of the shock-wave, which is especially probable in the process of formation of the shock-wave (in the pre-detonation period) and is more likely to take place near the walls of the tube than on its axis.

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Received
2.IV.1945.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

BIOCHEMISTRY

A METHOD FOR DIAGNOSTICATING GRAIN TOXIC AS A RESULT OF ITS HAVING OVERWINTERED IN THE FIELD

By W. L. KRETOWITCH and Z. G. SKRIPKINA

(Communicated by A. N. Bach, Member of the Academy, 26. II. 1945)

It has been established that human subjects feeding on such grain as had overwintered in the field under snow-cover are liable to suffer the attack of a severe disease called «septic angina», or «alimentary toxic aleucia».

This affection is a typical alimentary toxicosis caused by some poisonous substances that develop in a grain allowed to stay over winter in the field under snow. These substances seem to have a specifically toxic effect upon the tissues of the marrow, causing a strong inhibition of leucopoesis and, as a result of this phenomenon, strong leucopenia at early stage of the disease, and septic, necrotic angina at its later stage.

As a result of medical expeditions and researches it has been established that overwintered poisonous grain can be distinguished from normal grain after the effect the oil, isolated from it with the aid of fat-solvents, exerts upon depiled skin of white rabbits (reddening, profound oedema and necrosis of the tissues). Unlike this, oil extracted from normal grain has been found to produce no effect whatever upon the skin of rabbits. This observation was laid at the basis of biological method for diagnosticating toxic overwintered grain and products prepared with such grain. This method proved to be rather cumbersome, however, requiring no less than 7—8 days before a definite conclusion can be reached.

The object of the present investigation was therefore to investigate the poisonous grain both biologically and biochemically, and to develop a new, simpler method for the discrimination between normal and poisonous grain.

The samples of grain which served for our work had been confiscated as undoubtedly poisonous food-products by Professor P. G. Sergiev, Professor V. V. Efremov and Doctor S. B. Shimshelevich at various regions of the USSR from such households, where cases of septic angina had been recorded. Numerous samples of non-poisonous millet from the various parts of this country were put at our disposal by M. K. Firsova, a distinguished specialist in this crop, to whom our sincere thanks are due for many a valuable advice given in the course of this work.

Our experiments embraced millet, wheat and rye; when working with millet, carefully extracted kernels were made to germinate. As testified by many physicians, millet produces a particularly harmful effect.

There is absolutely no difference between the exterior appearance of poisonous or normal grain. The absolute weight of both is also quite similar.

But the energy of growth and the germination power are considerably reduced in poisonous grain. In the samples investigated by us they have been found to fluctuate within 26—40 and 43—59 per cent, respectively. But low germinating can by no means, however, be interpreted as a character testifying to the

toxicity of a given sample of millet; cases are quite possible where under unfavourable conditions of harvesting or storage, the grain may lose its viability without becoming poisonous. Thus, for example, a sample of pure-bred millet of good quality after being subjected by us to experimental self-heating was found to have lost all of its germinating power because of the profuse growth of fungi and bacteria; when tried by the method of skin-test, however, it was found to possess no toxic properties. Likewise inoffensive was found, when tested by the same method, a sample of millet which, because of unfavourable harvesting conditions, had been badly affected by various fungi and had therefore lost its viability. If the external appearance of the millet grains gives no criterion to discriminate between poisonous or normal grain, striking difference between the kernels of poisonous and normal millet is observed when the kernels are extracted with the aid of a Lebedev millet-sheller or by hand, with the aid of a needle. It has been found that the kernels obtained from samples of poisonous millet may be divided into the following three fractions: 1) kernels of almost normal appearance, though looking somewhat whitish, as it were discoloured when compared to the normal ones; 2) grey and brown kernels, and 3) kernels absolutely black in colour, as it were stained black from outside.

The percentage of grains with ear scales and containing kernels of different types is given in Table 1.

Table 1

Percentage of Grains Containing Kernels of Different Types

No. of sample of toxic millet	With ear scales	With light kernels	With dark kernels	With black kernels
1	9	61.5	22.5	16
2	6.5	60.5	25.5	14
3	7	63.5	2.9	7.5
4	8.5	59	31	10
5	4.5	56	28	16
6	7.5	73.5	22.5	4
7	6	81	15.5	3.5
8	10	61	22.5	16.5
9	5.5	67	29.5	3.5
10	4	67.5	29	3.5
11	0	62	32	6
12	0	85.5	13.5	1
13	26.5	79	17.5	3.5
14	2.5	81	15	4
15	5	69.5	24.5	6
16	9.5	57	36	7
17	0	86	9	5
18	4	84.5	9	6.5
19	5	68.5	28.5	3
20	4	86.5	12.5	1
21	50	78	10.5	2.5
22	8	80.5	18	1.5

It is thus evident that the samples of millet investigated by us contain a large number of grains badly affected by microorganisms. Microbiological analysis went fully to confirm this conclusion. The determination of the absolute weight of 1000 kernels has shown that the weight of darkened or black kernels is smaller than that of kernels of normal, light colour. In some of the samples this reduction in the absolute weight of kernels attains 17.5 per cent.

It should also be pointed out that the black kernels are very brittle.

All the samples of toxic millet investigated by us, irrespective of the year and locality of harvest, have been found to contain considerable numbers of grains whose kernels had the characteristic appearance and looked as it were stained black.

For the sake of comparison we repeated the analysis on 100 samples of pure-bred millet of good quality, harvested at different years and in different regions of the USSR. Only five of these samples have been found to include grains containing black kernels at a ratio not exceeding 0.5 per cent. These samples have also been found, however, to include 2—5 per cent of grains with darkened (grey, brown, dark brown) kernels. One of the samples of pure-bred millet was found to contain up to 21 per cent of grains with darkened kernels and 3 per cent of grains with black kernels. This sample could, consequently, be referred to samples of toxic millet.

Summing up the above-said, we feel justified to state that the presence in a millet-sample of 0.5 per cent of grains with black kernels, and of a considerable quantity (above 5 per cent) of grains with darkened—grey, brown, dark brown—kernels, makes one question the good quality of the sample and its adequateness as a food-product.

In such doubtful cases the definitive judgment should be based upon the results of the skin-test.

The samples of toxic wheat and rye were likewise investigated by the method described above.

The grain was differentiated into three fractions: 1) grains of normal appearance; 2) shrivelled, meagre and whitish grains; 3) shrivelled and meagre grains, blackened either all over the surface, or by spots; especially characteristic is the blackening of the embryo, which is either superficial, or may easily be revealed by cutting the embryo in two. We thus see that, like in millet, the formation of a black pigment is the most characteristic symptom of overwintered, toxic grain in wheat and rye.

In some of the samples of wheat and rye the percentage of blackened grains was as high as 52—56. The determination of the absolute weight of the poisonous grains of wheat and rye according to the different fractions has shown that the black grains were not so heavy as the grains of normal appearance.

In so far as the degree of energy of growth and germinating power of a grain are the fundamental indices of its affection by microorganisms, a comparison of the various fractions of toxic grain for this index may characterize the extent to which these fractions are injured by microorganisms.

Part of the results of our determinations are given in Table 2.

Table 2

Germinating Power of Toxic Grain According to Fractions

(in per cent)

	Fraction*	Germination, %	
		in three days	in seven days
Rye	1	53	60
	2	10	12
	3	6	8
Wheat	1	23	50
	2	5	29
	3	0	4
Millet	1	50	66
	2	15	16
	3	0	0

* 1, grains of kernels of normal appearance; 2, intermediate fraction of grains or kernels; 3, black grains or kernels.

From the data tabulated above it may clearly be concluded that both the energy of growth and germinating capacity are very low even in such grains of wheat and rye and kernels of millet whose outward appearance is quite normal. This should be taken to mean that they are also affected by microorganisms. When made to germinate, they give an abundant growth of fungi and bacteria. This observation is in full agreement with the data of physiological tests of the oil obtained from such normally looking grains and kernels: such oil has likewise been found toxic when tested by the method of skin-reaction described above (all the physiological tests were carried out at the laboratory of Professor V. V. Efremov). The oil produced from the intermediate fraction also proved poisonous, while the energy of growth and germinating power of the grains within this fraction were appreciably below those of the grains and kernels of normal appearance. The fraction of blackened grains and kernels was found to be practically deprived of any germinating power.

It is of interest to note in this place that the oil obtained from black grains and kernels called forth a very intense reaction on the part of the skin of the rabbits tested; the reaction was much more intense than in the case of oil obtained with the two other fractions of the same sample. It may thus be assumed that endowed with particular toxicity are those samples of grain which contain a high percentage of black grains, or of grains with black kernels, which are badly affected by microorganisms.

The fact observed by us, *viz.* that samples of overwintered grain do necessarily comprise a considerable quantity of black grains (or, in millet, of grains containing kernels with a black surface), has been laid by us at the basis of a simple and rapid method for determining overwintered poisonous grain, the use of which for food-purposes is responsible for the disease called the «septic angina».

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Received
26.II.1945.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

GEOCHEMISTRY

ON THE REGULARITY OF DISTRIBUTION OF COAL ASH COMPONENTS THROUGHOUT THE AREA OF THE PECHORA BASIN, AND THE DEGREE OF WEATHERING OF COAL FROM DATA OF TECHNICAL ANALYSIS AND ANALYSIS OF COAL ASH

By G. N. EGOROV

(Communtcated by V. A. Obruchev, Member of the Academy, 8. VII. 1944)

A. 1. Analyses of ash coals from Lower Permian and Upper Permian deposits show a uniformity of composition of ash for coals from deposits located within the area of the Pechora coal basin. Table 1 shows the composition of ash of coals according to deposits.

Table 1

Name of deposits	SiO_2	Al_2O_3	Fe_2O_3	CaO	MgO	SO_3
Lower Permian						
Silovaya	68.12	23.50	5.45	2.70	0.58	1.75
Khalmer-Yu	62.43	21.08	7.86	3.54	1.60	2.84
Upper Syr-Yaga	59.91	23.02	10.46	3.22	1.13	1.52
Eletz	61.80	19.90	10.02	2.41	0.99	2.6
Vorkuta	63.24	23.20	10.00	2.50	1.01	0.78
Inta	62.70	21.50	8.80	2.50	0.36	3.65
Talbey	60.36	23.80	5.60	5.00	1.08	3,52
Upper Permian						
Khalmer-Yu	61.42	27.22	5.30	3.40	1.51	1.54
Central	66.20	18.86	9.44	2.50	0.72	1.96
Yus-Yol	60.00	19.49	11.96	4.75	1.89	1.95
Talbey	66.77	19.85	6.80	2.60	1.16	2.41

2. From Table 1 it may be seen that the SiO_2 content in the ash of coals is regularly distributed over the area of the basin. Thus, for instance, the coals from deposits adjoining the western slope of the Polar Urals contain a greater amount

of SiO_2 than those located at a certain distance westward from it.This is true both of the Upper and of the Lower Permian coals (see the Talbey deposit). A similar change is observed also for the components of ash—Al_2O_3, Fe_2O_3, MgO and SO_3 (Fig. 1), with the exception of CaO, the content of which increases in the components of ash from East to West (Fig. 1).

Fig. 1. Regular variation of the ash components. Scale 1 : 1 000 000. Ash components: *I*, SiO_2; *II*, Al_2O_3; *III*, Fe_2O_3; *IV*, CaO. *1*, Silovaya; *2*, Khalmer-Yu, *3*, Central; *4—5*, Upper Syr-Yaga; *6*, Vorkuta; *7*, Eletz; *8*, Talbey; *9*, Inta; *10*, Yus-Yol. Variation in the content of ash components from East to West: SiO_2→decrease, Al_2O_3→ decrease, Fe_2O_3 → decrease, CaO←increase, MgO→decrease from 2.00 to 0.6%, SO_3→decrease from 4.00 to 0.9%

3. The existence of a regular change, a decrease or an increase in some component of ash within the area of the basin, may be connected with the palaeogeographical conditions of sedimentation and formation of coals.

4. No direct relation between the composition of a definite grade of coals and the composition of ash could yet be established, but the treatment of a large number of analyses of ash from most of the deposits of the basin is likely to disclose such a relation.

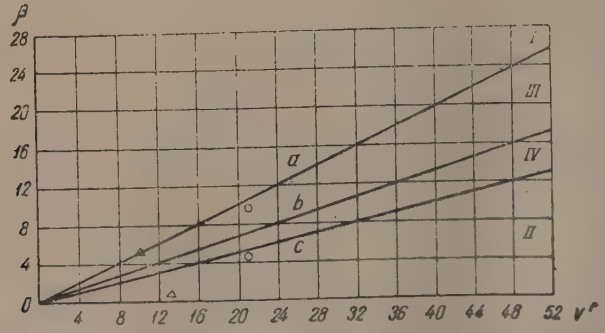


Fig. 2. Indicator of weathering of coal β: *I*, zone of non-weathered coal; *II*, zone of weathered coal; *III*, zone of the lower boundary of weathering of coals of grades from T to K. Lower boundary of CaO content of coals: *a*, of grades T—K; *b*, of grades К—Г; *c*, o grades ПЖ—Д

B. 5. The relation between the composition of the ash of coals and the degree of weathering is determined by the ratio of the amount of volatiles to the combustible mass V^c to the amount of CaO in the ash of coals. This ratio $\beta = V^c/CaO$ is called the indicator of the weathering of the coal (Fig. 2).

6. The numerous analyses of coals from the Pechora basin show that such a

relation is observed nearly for all the grades of coals. It is directly connected with the depth of the zone of permanently frozen ground. Thus, for instance, the analyses of coals from southern deposits of the basin of Yus-Yol, Inta point to a deep zone of weathering, while the analyses of coals from the northern deposits—Silovaya, Khalmer-Yu, Upper Syr-Yaga, Eletz—indicate an insignificant depth of the zone of physical and chemical weathering of coals.

7. In connexion with the behaviour of the zone of weathering, a distinct relation is revealed between the composition of the ash and the volatiles in the weathered coals. It is expressed in a higher content of volatiles to the combustible mass, and a higher content of the CaO component of ash in weathered coals. Table 2 contains data illustrating this relation.

Table 2

Upper Syr-Yaga Deposit

Bed	Place of sampling with regard to the river level	Quantity of volatiles V^c, %	Quantity of CaO, %	$\beta = \frac{V^c}{CaO}$
Bed 1	Open cut 1, depth 2 m.	10.13	2.60	5.10
»	Open cut 3, depth 0.5m.	10.26	8.25	1.28
»	Test-pit 3, depth +1.3 m above river level	13.13	26.02	0.50
Bed 7	Ditch 6, depth 0.5 m. .	11.19	3.20	4.50
»	Test-pit «Б», depth+1.3 m above river level	9.16	1.75	5.25

8. An examination of Table 2 will show that the lower limit of weathering of coals of grade T V^c=10 per cent (Upper Syr-Yaga deposit); β=3.50. The normal CaO content in non-weathered lean coals should be equal to 2—3 per cent, and for the ПЖ—Д grades from 3 to 4 per cent. This is borne out by a large number of coal ash analyses.

Grades from T to К

Khalmer-Yu	deposit	CaO	from	2.20 to	2.80	%
Silovaya	»	»	»		2.70	»
Upper Syr-Yaga	»	»	»	1.75 to	3.07	»
Eletz	»	»	»	2.00 to	2.85	»

Grades ПЖ—Д

Khalmer-Yu	deposit	CaO	from	2.80 to	4,20	%
Central	»	»	»	2.50		»
Yu-Yol	»	»	»	3.10 to	4.75	»
Inta	»	»	»	5.50 (weathered coal)		

9. Using data of analyses of coals, in particular V^c and CaO, and with the aid of the «Graph of the indicator of weathering of coals β», it is possible to determine the degree of weathering of coals and the reliabiity of a given analysis of a coal bed from the standpoint of its weathering.

Thus, for instance:

I. An analysis was made of coal of the К grade coking coals from a test-pit, from a depth of 1.5 m below the river level. Bed XII—Khalmer-Yu deposit: V^c=20.90 per cent; CaO=2.20 per cent; β=9.50.

Plotting the data on a graph (Fig. 2): along the abscissa, the amount of volatiles, and along the ordinates, β=9.50 is plotted

An examination of the graph will show that the analysis fell to zone *III*, that of the lower boundary of weathering of coals of the T—K grade, *i. e.* the analysis may be relied upon and considered significant for evaluating the quality of coals.

Bed XIX—Khalmer-Yu deposit. $V^c = 20.90$ per cent; CaO = 5.00 per cent; $\beta = 4.20$. Superimposing the value of β on the graph shows that it is located in zone *II*, that of weathered coal (Fig. 2).

II. Coals of the T grade are lean.

Bed 1—Upper Syr-Yaga deposit. Data of analysis: test-pit 3, $V^c = 13.13$ per cent; CaO = 26.05 per cent; $\beta = 0.5$. On the graph the analysis will fit within the zone *II*, that of weathered coal.

Bed 1—open-cut 1. Sample taken at a depth of 2.0 m below the river level. $V^c = 10.13$ per cent; CaO = 2.60 per cent; $\beta = 5.10$. On the graph, the analysis will be located in zone *I*, that of non-weathered coal.

C. 10. The analyses of the ash of coals from Upper Permian and Lower Parmian deposits show that:

1) The amount of ash in a coal does not depend on the composition of the ash, and this composition does not determine the degree of ash content of coals.

2) The composition of ash in coals is entirely dependent on the composition of the enclosing rocks of the coal-bearing deposits—polymikt sandstones, aleurolites, *etc.* This is evidenced by chemical analysis of the enclosing rocks of the coal-bearing deposits of the Lower Permian (Table 3).

Table 3

Rock	SiO_2	Al_2O_3	Fe_2O_3	CaO	MgO	SO_3	Loss on ignition
Sandstone polymikt, exposure 9, Bokovaya rivulet	60.00	26.33	6.67	5.00	2.22	—	
Aleurolite, exposure 1, Dosadny river	62.29	14.24	10.76	4.00	1.10	1.56	5.52
Coaly argillite, exposure 44, Syr-Yaga river . .	59.60	17.96	7.09	3.90	1.20	0.95	9.52
Argillite, pitch No. 9, Bokovaya rivulet . . .	49.00	21.30	8.29	6.10	1.32	1.32	18.0

3) Hence it follows that the ash content of coals depends entirely on the partial ratio of the amount of the coal mass to the mineral admixtures in the coal bed.

4) The latter conclusion is one of great practical interst, since the high ash content of coals is the cause of an increase in the amount of volatiles in coals. This is evidenced by the analyses of coals from two beds of the coal-bearing series about 85 m thick, containing lean coals of the T grade (Table 4).

Table 4

Bed	A^g	V^c	SiO_2	Al_2O_3	Fe_2O_3	CaO	MgO	SO_3
Bed 1, Syr-Yaga river. .	8.34	10.13	57.56	24.60	12.40	2.00	1.44	1.86
Bed 0. 18 m, Syr-Yaga river	57.54	28.63	62.87	20.90	9.40	3.95	0.9	1.64

5) Consequently, when the ash content of coals is high, the grade composition of the coal cannot be accurately determined.

11. All the above data concerning the composition of the ash of coals, regular change in the composition of ash within the area of the basin, the relation between the composition of ash and the degree of weathering of coals, as well as the causes of the approximate uniformity of the composition of ash, point to the fact that throughout the Pechora basin, poorly investigated with regard to the quality of the coals of all grades, it is already possible to record:

a) regardless of the composition of coals, a decrease in the content in the ash of coals of such components as SiO_2, Al_2O_3, Fe_2O_3, MgO and SO_3, in a direction from East to West, and an increase in the quality of CaO in the ash of coals from East to West;

b) the composition of ash of weathered coals along with the technical composition of coals enables us to judge of the degree of the weathering of coals;

c) a qualitative characteristics of the composition of the ash of coals shows that the composition of ash depends entirely upon the composition of the enclosing rocks, and not on the amount of ash in the coal.

12. It is to be hoped that subsequent investigations of the composition of ash in coals of all the deposits of the basin, will further contribute to our knowledge of the processes of the formation of the basin, and to settle the question as to the technological properties of the ash of coals, their difficult fusibility, which is of great practical importance.

Received
8. VII. 1945.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

GEOLOGY

TIN MINERALS IN THE MAIKHURA SKARN DEPOSIT (GHISSAR RANGE)

By A. G. IVASHENTZOV

(Communicated by A. E. Fersman, Member of the Academy, 5. VIII. 1944)

Among the numerous minerals of the skarn deposits of Central Asia and of the Zeravshan-Ghissar mountainous country, in particular, the tin minerals—cassiterite and stannine—are the least known.

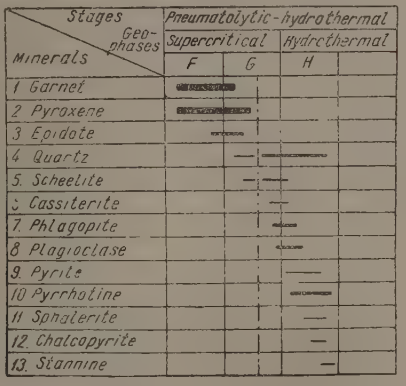
In the first time cassiterite and stannine were determined in appreciable amounts in 1933, in the ores of the Takfon skarn deposit, Zeravshan Range (1). Investigations carried out by the present author in 1939—42 also established the presence of cassiterite and stannine in the ores of the Maikhura skarn deposit, Ghissar Range (2). In 1939 I. G. Magakian recorded the presence of cassiterite in a number of small skarn ore occurrences on the southern slope of the Ghissar Range (3). The cassiterite and stannine of the Maikhura deposit are of particular interest on account of their intimate paragenesis with scheelite.

The Maikhura skarn tin-tungsten deposit is located in the upper reaches of the Maikhura, the right constituent of the Varzob River. The formation of skarn is confined to the area of contact of the Variscian intrusion of a granodioritic composition with a series of marmorized limestone within the metamorphosed limestone-shale formation of the Middle Palaeozoic. The principal skarn occurrence of a pyroxene-garnet composition is traceable for a length of over 300 m, having a thickness up to 25—30 m in the middle portion. In contradistinction from scheelite, cassiterite is absent from typical skarns of Maikhura, while it is found in appreciable amount in such sulphide-quartz ores only as are deposited along the fissures cutting the skarns. The sulphide-quartz ores are confined to steep-dipping zones of fracture inside the skarn deposit and to certain areas of the hanging wall of the latter, «screened» by marbles, to which the mineralization is not traceable. Some of the steep-dipping lenses of sulphide-quartz ores extend beyond the footwall of the skarn deposit into granodiorites.

The existence of a definite relation between the tin-tungsten mineralization in skarns and the fracturing is quite obvious. Aside from large fissures filled with sulphide-quartz tin-tungsten ores, the deposition of ores is also influenced by microfractures in skarns, along which in addition to quartz and sulphides, scheelite and cassiterite are also deposited.

The mineralogical composition of the sulphide-quartz ores varies greatly. They consist mainly of quartz (15 to 35 per cent), plagioclase (oligoclase) (15 to 25 per cent), phlogopite (5 to 20 per cent), sulphides: pyrrhotine (10 to 35 per cent), sphalerite (3 to 10 per cent), pyrite (1 to 2 per cent), *etc.*, as well as of appreciable amount of scheelite, cassiterite and a small quantity of chalcopyrite with stannine.

A paragenesis of scheelite and cassiterite with plagioclase and phlogopite is characteristic for the ores. Their structure is granular. In addition to the most important minerals enumerated above, apatite, chlorite, tourmaline, actinolite and some other minerals also take part in the composition of sulphide-quartz ores, usually in insignificant amounts. The sequence of formation and the paragenesis of the most important hypogene minerals of the Maikhura deposit (skarns and sulphide-quartz ores) is illustrated by the accompanying table.



Paragenesis of the most important hypogene minerals of the Maikhura depoits. Dotted line shows the time of manifestation in intra-ore fractures.

With regard to the time of formation one of the first minerals among the minerals of the sulphide-quartz ores is cassiterite.

Macroscopically, two varieties of cassiterite are distinguishable. The first variety is represented by large crystals of a short-columnar habit, up to 40—50 mm long and 20—30 mm across. The colour of the crystals is light-brown, sometimes greenish. This type of cassiterite is associated with plagioclase, phlogopite and scheelite. The second variety of cassiterite is represented by small, dark, almost black crystals of a columnar habit, sometimes twinned. The grains are up to 10 mm long and up to 1—2 mm across. Small grains, the size of which is but some fractions of a millimetre, are of most frequent occurrence. This variety of cassiterite is associated with sulphides (chiefly with sphalerite and pyrrhotine), plagioclase and phlogopite. The lustre of cassiterite is adamantine in both varieties. In the crystals of cassiterite of the second variety prism and pyramid faces are well developed. Under the microscope no essential difference between the two above described varieties of cassiterite is revealed.

In thin-sections the cassiterite grains are either colourless or slightly coloured in brownish shades. In the coloured grains a slight pleochroism and zoning are observed. Some of the cassiterite crystals (according to the data of E. F. Ziv) in the conoscope show the pattern of a biaxial mineral with a very small optic angle.

In polished sections cassiterite is distinguished by its high relief and its being poorly adapted to grinding. In oblique light it is translucent, showing a dark-brown colour. A considerable part of cassiterite grains are highly fractured. The fractures are filled with sulphides: pyrrhotine, sphalerite and chalcopyrite.

The chemical composition of the pure cassiterite of Maikhura (according to data by E. F. Ziv) is as follows:

SnO_2	97.9%	Al_2O_3	0.11%	MnO	traces
TiO_2	0.25%	SiO_2	0.58%	CaO	0.07%
FeO	0.22%	Nb + Ta	0.08%	Fe_2O_3	—
MgO	0.04%	ZrO_2	0.06%	S	0.18%
					99.49%

By the above described typomorphic properties the cassiterite from Maikhura fits the most closely, according to Fersman's scheme (4), the minerals of the pneumatolytic supercritical geophase G and the hydrothermal geophase H.

The second tin mineral in the Maikhura deposit, which has not found any practical application, is a sulphosalt of tin—stannine (Cu_2S. FeS. SnS_2). This mineral, determinable only microscopically (when examining polished sections), occurs in sulphide-quartz ores in the shape of very small inclusions—from 0.005 to 0.3 mm, generally among grains of chalcopyrite and sphalerite, but also as very thin fringes around grains of chalcopyrite, or in its fields as minute disseminated grains approaching in size an emulsion impregnation. As to time of its formation, stannine seems to have appeared later than all of the rest of sulphides, that is its segregation terminated the ore hydrothermal process of mineral formation.

The presence of cassiterite in the black sands from river deposits almost throughout the area of the Zarevshan-Ghissar mountainous country, along with its high concentration in the Maikhura and Takfon deposits, points to the necessity of a thorough search for new primary deposits within the area named.

Received
5.VIII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

MINERALOGY

AN UNDESCRIBED IRON METEORITE EL QOSEIR

By E. L. KRINOV

(Communicated by W.I.Vernadsky, Member of the Academy, 30. VII. 1944)

Among the samples that make up the meteorite collection of the Academy of Sciences of the USSR there is one iron meteorite which was neither described in literature, nor has been listed in any of the meteorite catalogues published to this date.

This meteorite has attracted the attention of the present author while he was engaged in systematizing the collection upon its re-evacuation. He has succeeded in finding out the following circumstances bearing upon this meteorite.

The meteorite was purchased by the Academy of Sciences on November 17, 1925, from a boatswain of the Soviet merchant shipping, Peter Churilov. In the archives of the Committee for Meteorite Research of the Academy of Sciences of the USSR a deed is preserved, which was signed at the above date by L. A. Kulik in the presence of P. N. Churilov, on the one hand, and of D. O. Mokhnach and G. Ognev who at the time belonged to the personnel of the Meteorite Department of the Mineralogical Institute of the Academy of Sciences of the USSR, on the other. This document testifies to the conveyance by Churilov to the Academy of the meteorite in question, describing in short the circumstances under which it had been found. The meteorite was valued at 96 roubles and 20 copeks, which made 4 copeks per gram, the total weight of the meteorite being 2405 g. This sum was payed to Churilov as a prize for his finding, preserving and delivering the meteorite to the Academy. The relatively high prize was agreed upon on account of «the integrity of the meteorite, which represents an individual body, its small weight (less than 10 000 g), its shape and, lastly, the information as to the place of its finding».

The circumstances under which the meteorite has been found are as follows. On the 16th of March, 1921, P. N. Churilov arrived on board the ship «Nizhny Novgorod» at a small Egyptian town, El Qoseir, lying on the Egyptian coast of the Red Sea ($\varphi = 26°17'$ of northern latitude, and $\lambda = 34°15'$ east to Greenwich*).

The shore-line, Churilov said (see the document cited above), is entirely made up there of coral reefs; farther come dunes and farther still, alabaster terraces are rising, which are occasionally broken by narrow corridor-like creeks leading to closed bays, no more than 1 km long.

While the ship took in her cargo at El Qoseir, which occurred between the 16th and the 29th of March, Churilov was boating along the shore. It was during

* The English transcription of the name of this town has been adopted from map No. 79 attached to The Encyclopaedia Britannica, London—New-York, 1930, vol. XXIV (this name is missing in the index to the corresponding volume). The Russian transcription has been taken from the Geographical Atlas for Red Army Officers, map No. 30, published by the General Staff of the Red Army, in 1938, in Moscow. The coordinates of the town El Qoseir have been determined by the author according to the same map.

one of his small excursions to the bay, to a distance of about 15 km south from the town El Qoseir at one of the terraces referred to above, that the meteorite was found. It was almost entirely exposed at the surface of the ground. Churilov was unable to gather any information as to the time and circumstances of the fall of the meteorite. . .

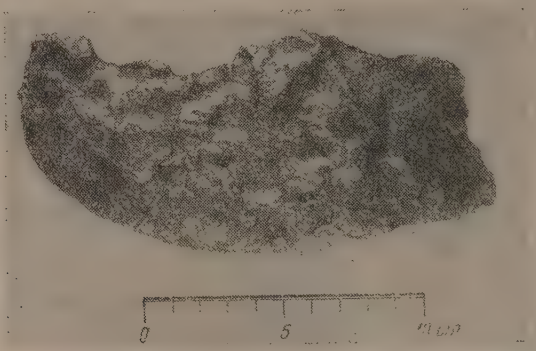

Fig. 1.

As has been said above, the original weight of the meteorite was found to be 2405 g; it was entered into the recording book of the Meteorite Committee under number 135.

On the 17th of August, 1932, the following parts were clipped off the meteorite:

No. 649, a convex portion	31.60 g
No. 650, a section	18.60 »
No. 651, a small angular fragment	7.75 »
Sawdust collected in two test-tubes	6.45 »
	2.51 »
Total	66.91 g

Now the weight of the main bulk of the meteorite was evaluated by the author at 2320.7 g; if summed up with the weight of the portions severed from it, a value

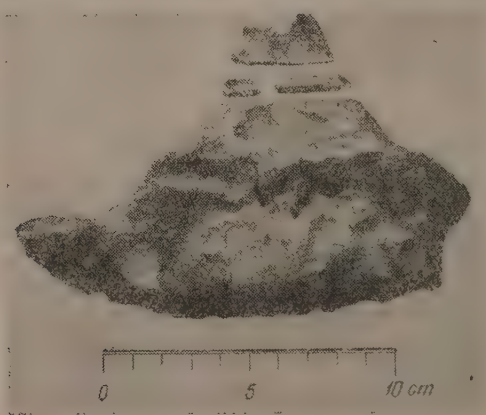

Fig. 2.

is obtained which coincides with that of the original weight of the meteorite, which has been cited in the beginning of this note. Thus, the loss on sawing made up 17.39 g. The main bulk of the meteorite together with the pieces that had been sawn off, save for the sawdust, are presently preserved at the Meteorite Collection

of the Academy of Sciences of the USSR. The sawdust, 8.96 g in weight, was taken in 1932 by V. Smirnov who analysed it for nickel. The analysis revealed a «high» nickel content; no detailed determination of the chemical composition of the meteorite was made, however. At the present author's request, D. P. Maliuga, from the W. I. Vernadsky Laboratory for Geochemical Problems at the Academy of Sciences of the USSR, undertook in May, 1944, to determine by means of a polarographic method the content of nickel, cobalt and copper in the El Qoseir meteorite; the following results have been obtained: Ni=15.17 p. c., Co=0.92 p. c. and Cu=0.32 p. c. The meteorite under description should thus be classed among the group of ataxites. Etching of the polished surface of sample No. 650 with 1 per cent solution of nitric acid went to confirm this diagnosis. During the etching test made in May 1944 by the author, neither Widmannstättian figures nor Neumann lines could be obtained.

The specific gravity of the meteorite was determined pycnometrically by T. A. Burova, of the Mineralogical Institute of the Academy of Sciences of the USSR, in August, 1932, and was found to be 7.783.

The dimensions of the meteorite are $7 \times 10 \times 15$ cm; at the first sight it appears to be an irregular shape. Upon closer examination one soon becomes aware that its elements are oriented relative to the direction of its movement through the Earth's atmosphere, prior to the loss of cosmic velocity (before it reached the «zone of retardation»). The anterior part of the meteorite consists of an arched edge formed by the crossing of its two largest planes (Fig. 1). This edge bears distinct traces of the action of the Earth's atmosphere: it shows neither projections nor depressions which are usually observed on the surface of meteorites. One of two planes referred to above has also been strongly affected by the atmosphere. Unlike this, the other plane shows a sharply pronounced regmaglyptic relief (Fig. 2). It should be pointed out that faces of all the meteorite under description display a regmaglyptic pattern which may serve to a certain degree as evidence of its relatively recent fall, for otherwise its surface would have suffered considerably, as it is usually the case with other iron meteorites. The dryness of the climate which prevails at the locality where the meteorite was found, should also be taken into consideration.

Thus, acccording to its shape the meteorite under description should be classed among partly oriented meteorites.

The meteorite has been named El Qoseir after the town near to the place of its finding. Both in the register and the archives this name was first spelled «Kosseir», or «Косеир» in Russian transcription.

The Committee for Meteorite Research.
Academy of Sciences of the USSR.

Received
30.VII.1945.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

HYDROGEOLOGY

VISCOSITY OF UNDERGROUND WATERS OF THE PERMIAN AND CARBONIFEROUS OF THE POLAZNA-KRASNOKAMSK ANTICLINE

By A. M. KUZNETZOV

(Communicated by W. I. Vernadsky, Member of the Academy, 11. VIII. 1944)

The coefficient of internal friction of underground waters and brines, as well as their density, is always taken into account in hydrogeological engineering designs.

The values of d and η, characteristic of edge waters, are used when correcting test pumping-out and forcing-in. A knowledge of the viscosity of the crust waters is of interest also for an understanding of their migration and geochemistry.

The measurement of the viscosity of Permian and Carboniferous waters of the Near-Kama area, whose chemical composition was discussed in a previous paper (1), had been carried out by the standard method in the Ubbehlode—Golde capillary viscosimeter with a manostat (2) at temperatures of 5, 10, 15, 20 and 25 C. Apart from some changes of design, the manostat consisted of two vessels, one 15 and other 50 l in capacity, connected in series to a vacuum pump by means of rubber tubing and three-way vacuum stop-cocks, and to a viscosimeter and water manometer. Such a connexion makes it possible to secure a smooth course of discharge in the system and to maintain the pressure during the outflow of the liquid at the desired level. The viscosimeter was placed in a water electrically heated thermostat with 10 l capacity, in which the temperature varied during the time of measurement by no more than 0.1°C. The time of outflow was usually measured under a pressure of 200 and 300 mm of water and was always found to exceed 60 seconds.

The viscosimeter had chambers of 19 cm^3, length of capillary 15.0 cm. and a diameter of 0.044 cm. The value of the Reynolds criterion, even for an outflow of distilled water at 25°C, fully satisfied to the condition of laminar flow. In order to calculate η the average values of three outflows with pressures of 200 and 300 mm were used, and the averages of two values were ultimately calculated.

The values of the viscosity of the waters investigated are given in the table.

The viscosity of the brines of the Carboniferous is characterized by closely similar values, differing from those of the Permian waters. A plot illustrating the relation between viscosity and temperature, which is omitted here, shows a smooth run of the curves; it can be seen that the viscosity of the mineral waters, brines in particular, decreases with a rise of temperature at a more rapid rate than that of pure water.

The fluidity $1\,\eta$ is in a linear relation to temperature and can be expressed by the equality $1/\eta = a_i + b_i t$, which makes possible to calculate the viscosity at any temperature within the range studied.

Region	Hole	Depth, m	Density at 20° C	Content of salts per 1000 g				Vicosity at t°C in centipoises				
				in g	ing-equiv.	$CaCl_2$	$MgCl_2$	5°	10°	15°	20°	25°
Krasnokamsk . .	25	94—360	1.027	38.0	1.274	—	—	1.571	1.380	1.219	1.096	0.992
U-Kachka	2	211—480	1.056	73.2	2.480	—	1.2	1.652	1.453	1.281	1.153	—
»	19	120—300	1.077	105.4	3.592	0.2	3.4	1.757	1.534	1.361	1.221	1.097
Nytva	21	232—240.5	1.163	207.4	7.220	27.0	13.3	2.416	2.131	1.882	1.675	1.503
Krasnokamsk . .	11	720	1.179	225.9	7.930	13.1	4.8	2.578	2.244	1.986	1.756	1.596
Polazna	2	1057—1095	1.180	221.8	7.827	52.3	22.2	2.692	2.352	2.072	1.842	1.640
Severokamsk . . .	5	1030—1050	1.173	217.0	7.691	53.8	25.1	2.673	2.336	2.065	1.835	1.634
Krasnokamsk . .	62	987—990	1.172	217.1	7.590	43.7	14.7	2.579	2.245	1.990	1.758	1.598
»	66	1054—1060	1.172	216.8	7.576	44.2	13.9	2.629	2.266	2.017	1.790	1.619
»	82	1184—1188	1.174	220.4	7.698	39.4	16.6	2.637	2.298	2.031	1.810	1.620
Severokamsk. . .	10	1347—1348	1.167	210.8	7.360	42.0	13.1	2.511	2.200	1.944	1.728	1.554
»	4	1324—1325.5	1.180	222.1	7.843	44.8	13.6	2.666	2.333	2.048	1.828	1.632
Krasnokamsk . .	82	1252—1252.7	1.177	224.5	7.842	43.3	13.6	2.627	2.284	2.020	1.799	1.620
»	104	1243—1244.3	1.174	221.5	7.724	42.4	13.6	2.620	2.280	2.016	1.794	1.616
Severokamsk . . .	10	1370—1385.5	1.176	220.3	7.709	42.7	15.3	2.578	2.258	1.998	1.770	1.597

The coefficients a and b vary from 0.545 and 0.0181 (water No. 1) to 0.310 and 0.0115 (water No. 6), respectively.

Viscosity does not increase uniformly with an increase in the density and mineralization of waters. Brines of a similar density, but differing in the total contact of salts and their composition, differ in the values of η (Nos. 8 and 9; 6 and 12; 10 and 14).

The viscosity of the waters of the Permian is lower than those of the waters of the Carboniferous, and increases along with the mineralization, which coincides with the pitch of the water-enclosing rocks of the Permian following the anticline*.

The chemical composition of the brines shows that the increase in viscosity does not only depend on the salt concentration, but also on the amount of $MgCl_2$ and $CaCl_2$.

Waters containing high concentrations of these salts show also a high viscosity, which fact is accounted for by a greater viscosity of the solutions of these salts as compared to the solutions of NaCl of an equivalent concentration ([3]). This seems to be due to a greater association of Mg and Ca cations with water molecules and to the structural peculiarities of the bond, which becomes markedly pronounced along the decrease in temperature. Consequently, the viscosity of the natural waters containing differently hydrated ions, is not of the nature of a simple dependence on the total mineralization.

When determining the viscosity of the sulphide waters of the Permian, it has been found to increase according as H_2S is converted to colloidal sulphur.

It has been found that after the sulphide waters had been exposed to air during 2—3 days, the viscosity increased by $^1/_4$ to $^1/_5$ as compared to the value initially determined.

The values of the viscosity of the underground brines of the Near-Kama region very closely approach the values of the viscosity of the brine of the Crimean lakes ([4]). Under experimental conditions the viscosity of the brine seems to differ but very little from its viscosity under the conditions of the lake. This is not true of the viscosity of the waters in the Earth's crust, particularly of brines from great

* In paper ([1]) the following subscription is ommited under the figure: Structural map of the basement of the sulphate deposits (gypsum and anhydrites) of the Kungurian. The contour line + 100 at Polazna lies 10.8 m above the normal water level of the Kama.

depths, highly saturated with gas and subject to pressures amounting to hundreds of atmospheres. The increase in the number of atoms per unit volume of such systems, and the great association between the water molecules and salt and gas components, makes one assume that under such conditions the viscosity of the brine should be higher. In so far as mineralization of underground waters increases with the depth of occurrence, and sodium becomes gradually replaced by calcium and magnesium, the mobility of underground water decreases with depth, if the temperature increases at a slow rate.

The internal friction of the brines under natural conditions of occurrence cannot be sufficiently well characterized by viscosity measurements of degased waters; such data, however, are evidence of their low mobility in the Earth's crust, as previously suggested by Vernadsky([5]). The precise idea of the internal friction of brines in the Earth's crust, which is of great practical and scientific interest, can be formed by reproducing experimentally the natural conditions, that is, by saturating waters with corresponding gaseous mixtures, under definite pressures and temperatures.

Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

GENETICS

VIABILITY, RATE OF GROWTH AND MORPHOLOGY OF CARPS OF DIFFERENT GENOTYPES AS AFFECTED BY REARING CONDITIONS

By V. S. KIRPICHNIKOV

(Communicated by L. A. Orbeli, Member of the Academy, 5. VIII. 1944)

This paper deals with results of experiments carried out with four groups of carps at the Nikolsk Soviet fish-farm, Leningrad region, *i. e.* under climatic conditions very severe for carps.

In order simultaneously to obtain the four principal genetic groups of carp, differing from one another in two pairs of genes *A* and *a*, and *N* and *n* ([1,2]), and sharply differing in a large complex of characters ([3—7]), heterozygous scaly carps were crossed with nude carps. The heterozygous condition of the scaly parent was determined by two check-up crossings:

1. ♀ *Ann* (scaly) × ♂ *aann* (scattered)
2. ♀ *Ann* (scaly No.01) × ♂ *Ann* (scaly No.03)

All the three scaly parents tested proved to be heterozygous and to have a genetic structure *Aann*. Nude carps are always heterozygous. Consequently our crossing scheme and the table of the results expected have the following shape:

♀ *aaNn* (nude) × ♂ *Aann* (scaly)

♂ \ ♀	Gametes *aN*	Gametes *an*
Gametes *An*	*Aa Nn* linear	*Aann* scaly
Gametes *an*	*aa Nn* nude	*aann* scattered

The fry was reared in three different ponds. The third (third rearing) pond offered the most favourable conditions, then came the second (fifth new experimental) cold-water pond with poor food supply, and, lastly, the first (second spawning) pond, almost dried towards autumn, overstocked, poor in oxygen and abundant in free carbon dioxide.

In the autumn the ponds were fished; the first pond yielded 76.4 kg fish per ha; the second, 78.8 kg; and the third, 275.9 kg. Pronounced differences were recorded in respect to weight and morphological characteristics.

The differences by genotypes were also well marked. Let us discuss them in some details. In the following discussion the scaly and scattered fry will be referred to as fry of the first group, and the nude and linear—as the second group.

Character		1 pond			
		I*	II*	III*	IV*
Survival	number of fry caught	274	260	100	79
	ratio (%)	38.4	36.5	14.0	11.1
Weight	mean weight in g	1.24	1.03	0.87	0.89
	coefficient of variation (%)	72.8	84.3	42.2	40.4
Body length (mean) in cm		about 3.2 cm			
Exterior index $\frac{\text{length}}{\text{depth}}$		2.98 ± 0.02	2.90 ± 0.02	3.15 ± 0.02	3.09 ± 0.02
Percentage to body length	thickness of the body	19.36	19.59	19.13	19.59
	antedorsal space	—	—	—	—
	length of *D* fin	—	—	—	—
	greater girth	—	—	—	—
Mean number of rays	soft rays in *D* fin	—	—	—	—
	soft rays in *A* fin	—	—	—	—
	total number of rays in V fin	—	—	—	—
Mean number of gill-rackers on one gill-arch	on the inner margin	—	—	—	—
	on the outer margin	—	—	—	—
Mean number of pharyngeal teeth		—	—	—	—

* I, scaled; II, scattered; III, linear; IV, nude.

1. Survival and Segregation. In autumn all the ponds contained a majority of scaly fry, a smaller number of scattered, still less linear, and but few nude carps. This difference between the first and second group tends to increase greatly under unfavourable conditions. Thus, in the first pond carps of the first group were three times as numerous as those of the second group.

The quantitative relationships (differences for genes *N* and *n*) are:

1 pond. 534 : 170 or 2.98 : 1 (expected 1 : 1). Difference statistically significant
2 pond. 192 : 107 or 1.79 : 1 (» 1 : 1) » » »
3 pond. 1397 : 1222 or 1.14 : 1 (» 1 : 1) » » »

Let us compare the total number of scaly and linear fry with the number of scattered and nude (differences for genes *A* and *a*).

1 pond. 374 : 339 or 1.10 : 1 (1 : 1) Difference not significant
2 pond. 159 : 140 or 1.13 : 1 (1 : 1) » » »
3 pond. 1368 : 1251 or 1.09 : 1 (1 : 1) » » »

The coincidence of the results obtained for all three ponds permits of the conclusion that the scaly carps have a somewhat higher viability (of the order of 10 per cent) than the scattered, and that the linear carps are more viable than the nude ones.

The difference between the carps of the first and the second group in respect to viability was especially conspicuous in the first pond, characterized by unfavourable life conditions. In this pond the losses exceeded 90 per cent; in the second pond they amounted to 25 per cent, and in the third pond (not taking into account an accidental mortality at the beginning of the experiment) they were not higher that 5 — 10 per cent. This change in the relative survival of different genotypes

Experimental Ponds

2 pond				3 pond			
I*	II*	III*	IV*	I*	II*	III*	
104	88	55	52	**730**	667	638	584
34.8	29.4	18.4	17.4	**27.8**	25.4	24.4	22.4
11.67	10.82	9.17	9.25	**37.33**	35.98	32.91	29.96
46.3	39.8	28.8	34.3	28.7	**33.5**	27.7	20.2
7.39	7.30	6.77	6.50	**10.71**	10.41	10.34	10.16
2.78 ± 0.01	**2.77 ± 0.01**	2.85 ± 0.01	2.85 ± 0.01	2.77 ± 0.01	2.74 ± 0.01	2.86 ± 0.01	2.82 ± 0.0
21.23	21.01	20.59	20.81	**20.83**	20.47	19.99	20.39
50.57	50.80	50.65	**51,62**	50.39	50.49	50.51	**51.24**
34.91	**35.09**	34.28	34.29	35.75	**36.25**	35.07	35.89
88.71	**90.82**	88.21	89.06	**90.33**	**91.02**	87.37	88.87
19.04 ± 0.12	18,72 ± 0,13	15,28 ± 0,23	13,74 ± 0,34	**18.82 ± 0,10**	18.72 ± 0.12	16.40 ± 0.32	15,54 ± 0.33
4.96	**4.98**	3.30	3.06	**4.96**	**5.00**	3.82	3.56
8.87	8.87	8.24	7.99	**8.91**	8.68	8.76	8.47
30.80	30.20	26.22	26.00	**31.27**	30.79	27.49	27.55
24.44	24.40	20.10	19.90	**24.92**	24.29	19.37	19.37
9.20 ± 0.20	8.90 ± 0.28	7.22 ± 0.28	7.10 ± 0.28	9.22 ± 0.07	9.58 ± 0.09	7.63 ± 0.09	7.44 ± 0.10

is best explained by a greatly increased intensity of selection under unfavourable conditions of growth. To the expected segregation 1 : 1 : 1 : 1 satisfy to a certain extent only the relationships of the third and best pond.

2. Weight, Linear Sizes. Our estimates of mean weight and body length are of the same order as the estimates of viability.

3. Morphological Characters. As regards exterior indices, the differences between carps of the first and second group were manifest, being most pronounced for the first pond. Differences between scaly and scattered, and linear and nude, respectively, are not entirely statistically significant.

Carps of the second group (linear and nude) are known ([6–7]) to be characterized by strongly reduced fins and gills. This, too, was observed in our experiments. The reduction of fin-rays is especially pronounced in nude carps, their difference in this respect from linear carps is significant. It is of interest to note that the reduction of fins and gills tends to increase under unfavourable conditions of growth. There is also a slight difference in the number of gill-rackers between scaly and scattered carps. This increase of reduction is, however, somewhat questionable, for the number of gill-rackers is subject to age variations.

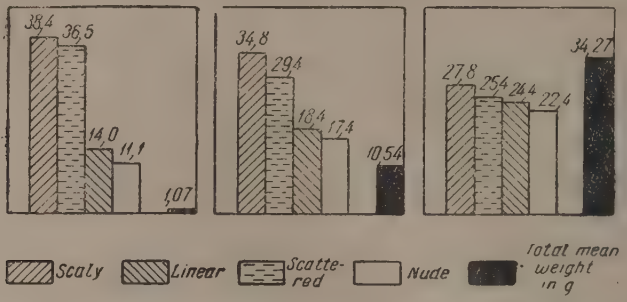
5. Pharyngeal Teeth. The teeth were counted and the differences observed in their number were checked on a material of other provenance and age. The two groups were found to differ in this respect also.

Carps of the first group are usually characterized by the teeth formula 1.1.3—3.1.1, rarely 1.1.3—3.1. The teeth are in three rows.

The usual teeth formula of the second group are 1.3—3.1, 1.3—3, rarely 1.1.3—3.1, 3—3, and 1.1.3—3. The teeth are always in two rows and reduced both in number and size. However, as a second generation of teeth takes place in the autumn ([8]), this increased reduction may be regarded as a possible lag in development due to a slowing down of the growth rate of linear and nude carps.

Conclusions. The essential morphological differences between carps within the groups investigated are retained under the severe climatic conditions of the Leningrad region.

The replacement of the gene *n* by its allele *N* brings about an abrupt change in the regularities of development during the earliest stages of embryogenesis. This change involves a large complex including such a constant character as the number and the arrangement of pharyngeal teeth. The changes attain a value surpassing many times the morphological differences between certain species of *Cyprinidae*. The general reducing effect of the gene *N* upon the development increases under conditions of poor food-supply and slow growth-rate. This points to a low resistance to external influences, a general «unstability» of structure in



Influence of rearing conditions on the segregation of fry (percentages)

linear and nude carps, and is also the cause of differences in rates of survival, especially pronounced under unfavourable conditions.

The replacement of the gene *A* by its allele *a* leads to less marked changes, though incontestable differences exist with respect to weight and linear growth, exterior and all meristic characters, especially between linear and nude carps. The gene *a* has but a slight reducing influence. No increase in reduction under unfavourable conditions could be observed, and no differences between the ponds in respect to relative viability. The segregation for genes *A* and *a* was found to be rather close to the expected one in all ponds.

The differences brought about by mutation offer an example of pronounced pleiotropism. Diverse physiological and morphological indices are subject to variation such as weight, linear growth, scales, dermal covering, body form, structure of fins, gills and pharyngeal teeth and, according to data as yet unchecked, certain internal organs, too. This should be taken to mean that the hereditary factors studied have probably a hormonal character.

The results of our experiments lead to the conclusion that linear and nude carps are entirely unsuitable for rearing in the northern zone of the USSR, on account of their lower exterior and weight indices, and especially of their low resistance to unfavourable environmental conditions.

The slight advantage that the scaly carps have over the scattered in respect to viability and rate of growth, though retained under any conditions, is somewhat handicapped by lower exterior indices. Scaly and scattered carps may almost equally be recommended for breeding in the North of the USSR.

Received
5.VIII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

GENETICS

A CONTRIBUTION TO THE STUDY OF THE INHERITANCE OF AWNEDNESS IN WHEAT

By G. B. MEDVEDEVA

(Communicated by A. A. Richter, Member of the Academy, 5. VIII. 1944)

The inheritance of awnedness has been studied in three crosses, in all of which the male parent was the awnless *Turgidum compositum* (English wheat with branched ears), and the female parents were: 1) *Melanopus* 069; 2) *Dicoccum rufum* (No. 25516 from VIR), and 3) wild wheat *Dicoccoides pseudoiordanicum* (No. 5201 from VIR), all three of them awned.

The aim was as follows. By rearing the F_1 and F_2 under appropriate conditions to give the development the desired trend, *i. e.*to bring out the possibility of obtaining directed variation in the hybrids. The F_1 plants were placed either into such conditions as would lead to the development (dominance) of the awnedness in the heterozygote, or to the dominance of the absence of the awns, *i. e.* to complete or partia (in a different manner in different crosses) suppression of the development of the awns. The common dominance of awnlessness in F_1 was observed in our crosses in the cases when the F_1 hybrids were grown under optimal cultural conditions.

The hybrids were sown in the end of April to hotbeds where they remained till maturation and harvest. The life space, soil, fertilizers and watering were such as to secure optimal natural conditions. The other portion of the F_1 hybrids was sown in the greenhouse in February, in a poor soil (mixture of clay with sand), no fertilizers were given, watering was scarce. As a result, dominance was modified with regard to many characters, awnedness included. Under these conditions the F_1 hybrids in all three combinations were awned.

The question arose, what will be the course of variation in awnedness in subsequent generations. Was the reversal of dominance in F_1 a mere phenotypical phenomenon, different reaction of the one and the same heterozygote to changing environmental conditions? In this case the F_2 should have displayed a uniform segregation for awnedness, irrespective of whether it came from awned or awnless F_1 plants, or perhaps the different manifestation of the character in the heterozygote (presence or absence of awns) was the result of the different formation of the hereditary basis under different environmental conditions. If it be so, the specific rearing conditions causing the character in F_1 should leave their mark in the second generation, even in the case of their absence at the time when the latter was reared; repeated action of the same conditions would increase the development of the same character.

The progeny of the awned F_1 was divided into two groups; one part of them was planted in the greenhouse: the other, in the hotbed. The progeny of the awnless F_1 was treated in the same manner. The same combinations of the environmental conditions which had created in the one case awned F_1, in the other—awnless one, were reproduced in rearing F_2. The F_2 was thus of four kinds: 1) from the awned F_1

under conditions causing awnedness; 2) from the awnless F_1 under the same conditions; 3) from awned F_1 under conditions causing awnlessness; 4) from awnless F_1 under the same conditions. The variant 3 perished (and so did partly the variant 4). The variants 1, 2 and partly 4 survived and could be employed for analysis and comparison.

The following points could be compared: a) the effect of different environmental conditions in F_1 under uniform conditions in F_2 (under conditions causing awnedness, *i. e.* in the greenhouse); b) the effect of different conditions in F_2 under uniform conditions in F_1; c) the F_2 from the awnless F_1 under the conditions causing awnlessness with the F_2 from the awned F_2, grown under conditions favouring awnedness.

Segregation for awnedness in F_2 displayed such a variety of the expression of awns that the hybrids could not be classified into awned and awnless ones. One had to take resort to measurements of the awns, to the construction of curves and to their comparison. In every instance only one spine was taken from each plant; one awn belonging to the middle spikelet was measured in each ear.

In the adjoining table are shown the lengths of the awns and their variation in the F_2 hybrids from the cross *Melanopus* 069 × *Turgidum compositum*.

Length of Awns in F_2 *Melanopus* 069 × *Turgidum compositum*

	Average in mm	Amplitude in mm	Number of plants measured
Melanopus 069 (♀)	86,1	70—110	100
F_1 and F_2 under «awned» conditions	19.92	0—100	84
F_1 under «awnless» conditions, F_2 under «awned» ones	18.62	0.90	285
F_1 and F_2 under «awnless» conditions	12.95	0.70	137
Turgidum compositum (♂)	0	—	100

The maximum length of the maternal awns was 110 mm. This length was attained by a fraction of the segregants only when the conditions which have created awns in F_1 were reproduced in F_2. The F_2 grown under the same conditions, but originating from an awnless F_1, gave the longest variants measuring only 90 mm, despite greater number of individuals (285 plants against 84). Still shorter awns had the segregants belonging to the extreme right part of the distribution of the F_2 grown during two generations under the conditions which «prevented» awnedness. Here the upper limit was 70 mm. The amplitude of the length of the awns in the F_2 proved to be in direct connexion with the conditions which were acting not only upon the last generation, but also upon the previous one. The extremes were shifted from 110 mm to 70 mm. Completely awnless plants in the F_2 were present in 3.57 per cent of individuals, when the greenhouse conditions were reproduced during two generations; when the hotbed conditions were reproduced during two generations, there occurred 19.71 per cent of awnless plants. Awned plants with awns not over 5 mm long were present in 42.84 per cent of cases in the former case, and in 68.62 per cent of cases in the latter case.

There was 59.5 per cent of awned plants with awns up to 10 mm long in the former case, and 77.38 per cent in the latter case. On the other hand, the higher classes possessing longer awns are represented by a lower proportion of plants with hotbed cultivation in F_1 and F_2 than with greenhouse cultivation.

In the two remaining crosses the directed variability of the character of awnedness was expressed with no lower distinction. In the cross *Dicoccum rufum* × *Turgidum compositum* the extreme F_2 variants, when cultivated in F_1 and F_2 under «awnless» conditions, attained only 55 mm, instead of 110 mm in the case of reproducing «awned» conditions during two generations. When the whole F_2 grown in the greenhouse is divided into two groups, the short-awned and long-awned ones, whatever the mode of division, the numerical prevalence would always belong in the short-

awned group to the F_2 originating from the awnless F_1, and in the long-awned group there will always prevail the descendants of the awned F_1. If one assumes the boundary at 5 mm, there will be 30.77 per cent of descendants of the awned F_1 in the short-awned group, and the descendants of the awnless F_1 will constitute 41.41 per cent, *i. e.* 11 per cent more. The same relation would persist if one takes 10, 15 or 20 mm limit. If one takes the long-awned group, *e. g.* having awns over 20 mm long (or 30, 40, 50, 60 mm), these groups would be more numerous in descendants of an awned F_1. When reared «toward awnlessness» during two generations, the hybrids displayed a still more expressed directed variation in the length of the awns: 78.33 per cent of the plants had awns not over 5 mm, and 95 per cent not over 20 mm, instead of 30.77 and 53.87 per cent when opposite («awned») conditions were applied during two generations.

In the third combination, *Dicoccoides* × *Turgidum compositum*, different rearing conditions of the F_1 also resulted in a considerable preponderance of awnless and short-awned forms in the offspring of the awnless F_1, and of the awned ones in the offspring of an awned F_1. There were 28.57 per cent of awns under 10 mm long in the offspring of the awnless F_1, and 9.8 per cent in the offspring of an awned. The figures for awns up to 20 mm long were 37.14 per cent, and 17.64 per cent respectively, *etc.*

On the other side, the offspring of the awned F_1 showed preponderance of the long-awned forms. Thus, there were 76.45 per cent of plants possessing awns above 30 mm, while the offspring of the awnless F_1 contained only 55.22 per cent of such plants, *etc.* When the conditions preventing the development of awns were reproduced during the two generations, the directed change in this character was expressed stronger still. In this case there were 70 per cent of plants with awns below 10 mm, instead of 9.8 per cent when «awned conditions» were reproduced in F_1 and F_2.

The plants having awns below 20 mm amounted to 30 and 17.64 per cent, respectively. The highest length of the awns with repeated «rearing of awnlessness» was 40 mm; with repeated «rearing of awnedness»—120 mm.

From the three crosses described here it follows that the character of awnedness dominates or is suppressed in F_1 according to the surrounding conditions which favour or suppress its expression. This effect persists in the next generation, and becomes stronger when the same conditions are reproduced.

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Received
5.VIII.1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

HYDROBIOLOGIE

LA DISTRIBUTION DE LA BIOMASSE DU PLANCTON DANS LA MER NOIRE

Par B. N. NIKITINE

(Présenté par I. I. Schmalhausen, de l'Académie, le 20. VII. 1944)

Une des questions les plus essentielles dans le problème de la productivité biologique des bassins aquatiques, est la question de la quantité et de la dynamique de la biomasse des organismes, habitant le bassin en question et, plus particulièrement, de la biomasse ou plancton.

Les particularités du régime hydrologique de la mer Noire, ayant pour cause l'absence d'oxygène dans les couches profondes, posent une limite à la distribution verticale des organismes aerobes vivants. Pour le plancton cette limite inférieure passe à 100—125 m dans les parties centrales et le long des côtes occidentales de la mer; dans la zone littorale des autres parties de la mer cette limite passe à 150—175 m, et, seulement, dans une région tres restreinte de la partie S—O et dans la partie occidentale des côtes d'Anatolie, cette limite s'abaisse jusqu'à 175—200 m et jusqu'à 200—225 m dans la région auprès du Bosphore ([1], [2], [3]).

Les données reçues pendant les expéditions de 1925—1928 (186 stations) nous permettent d'établir que la distribution de la biomasse du plancton, aussi bien horizontale que verticale, s'accomplie, dans la «zone vivante» de la mer Noire, très inégalement.

La distribution verticale de la biomasse du plancton par couches de 25 m peut être représentée comme suit (en mg/m^3): 0—25 m, 210 mg; 25—50 m, 147 mg; 50—75 m, 121 mg; 75—100 m, 84 mg; 100—125 m, 90 mg; 125—150 m, 54 mg; 150—175 m, 38 mg. La moyenne pour toutes les profondeurs est de 118 mg.

Parmi les agents qui influencent la distribution verticale de la biomasse du plancton, une importance particulière appartient à la température ([5]), ainsi qu'à la lumière, à la quantité de nourriture, *etc.*, mais dans les conditions de la mer Noire l'importance prédominante appartient à l'oxygène dont la quantité diminue sensiblement en dessous de 75 m et, en partie, au pH, la diminution duquel nous signale l'augmentation de la tension de l'acide carbonique libre ([6]).

La distribution horizontale de la biomasse du plancton dans la couche 0—100 m (fig. 1) nous apparaît sous un aspect suivant: auprès des côtes de la Crimée, des côtes occidentales, des côtes sud du Caucase, ainsi qu'auprès du Bosphore et dans quelques régions de la haute mer, la biomasse du plancton s'élève à un chiffre considérable (150—200 mg/m^3, même 200—300 mg/m^3), tandis que le long des côtes d'Anatolie et dans la plupart des régions de la haute mer ce chiffre diminue sensiblement (100—150 mg/m^3, même 50—100 mg/m^3).

Les agents principaux, conditionnant cette augmentation de la biomasse du plancton auprès des côtes de la Crimée, sont:1) l'afflux des eaux de la mer d'Azoff, riches en nitrates et phosphates, et 2) le renforcement de la circulation verticale des eaux, causée par les courants, et le phénomène d'un refoulement des eaux superfi-

cielles par le vent et leur remplacement par les eaux des couches profondes; auprès des côtes occidentales cette augmentation peut être expliquée, également, par l'afflux des eaux continentales (le Danube, le Dniepre, le Bug), contenant des sels biogènes, tandis qu'auprès des côtes d'Anatolie le courant amène des eaux apauvries d'éléments biogènes, ce qui explique le chiffre considérablement diminué de la biomasse du plancton. Dans les régions de la haute mer le chiffre de la biomasse du plancton augmente dans les endroits de convergences des courants où la circulation verticale devient plus intense et où les sels biogènes remontent vers les couches supérieures.

La distribution horizontale dans les couches 100—125 m nous montre que le chiffre de la biomasse du plancton, auprès des côtes de la Crimée, est relativement grand (50—100 m), ce qui trouve son explication dans une circulation verticale plus intense; auprès des côtes occidentales nous trouvons (contrairement à ce que nous avons trouvé dans la couche 0—100 m) ce chiffre considérablement diminué (25—50 mg/m³ et < de 25 mg/m³), en conséquence d'une circulation verticale

Fig. 1. La distribution horizontale de la biomasse du plancton dans la couche 0—100 m

affaiblie et d'une diminution de la qualité d'oxygène ayant pour cause la dessalaison des couches supérieures.

Auprès des côtes d'Anatolie nous trouvons dans ces mêmes couches un tableau tout à fait différent à celui des couches 0—100 m; le chiffre de la biomasse du plancton augmente (50—100 mg/m³) par suite de l'accroissement de la quantité d'oxygène, amené par le courant inférieur du Bosphore, lequel, en sortant du Bosphore, se dirige vers l'ouest et suit les côtes de l'Anatolie.

L'augmentation du chiffre de la biomasse du plancton entre les côtes de la Crimée et de l'Anatolie, ainsi que dans d'autres parties de la haute mer, est étroitement liée avec le renforcement de la circulation verticale, causé par les courants mêmes et les endroits de leur convergence.

Dans les couches 125—150 m et 150—175 m la distribution de la biomasse du plancton dépend entièrement de la différence dans le régime oxygène, cela veut dire, de la quantité plus ou moins grande d'oxygène, ce qui, à son tour, est conditionné par l'intensité, plus ou moins grande, de la circulation verticale des eaux.

De vastes régions de la mer sont, dans ces mêmes couches, entièrement privées de plancton (sur les cartes ces régions sont marquées par des tâches blanches), car sa limite inférieure, par suite d'absence d'oxygène, passe au-dessus de ces couches. Ces régions privées de plancton sont entourées de zones avec une quantité de biomasse insignifiante (< de 25 mg/m³), ce qui tient à un déficit d'oxygène.

Ce qui concerne la zone littorale et quelques régions de la haute mer (fig. 2), où la circulation verticale, sous l'influence des courants, devient plus intense et où nous pouvons signaler une amélioration du régime oxygène et un abaissement considérable de la limite inférieure du plancton, nous apercevons un accroissement du chiffre de la biomasse du plancton (25—50 mg/m³ et même 50—100 mg/m³).

Ainsi, la distribution horizontale de la biomasse du plancton dans les couches de profondeurs différentes dépend de deux agents essentiels: 1) de l'afflux des eaux continentales et des eaux de la mer d'Azoff, richent en sels biogènes, et 2) de l'intensité, plus ou moins grande, de la circulation verticale des eaux. Pour les couches supérieures les deux agents sont d'une importance égale, tandis que pour les couches profondes (en dessous de 100 m) le second joue un rôle prépondérant, car c'est lui qui conditionne le régime oxygène lequel, à son tour, conditionne la distribution de la biomasse du plancton dans les couches profondes de la mer.

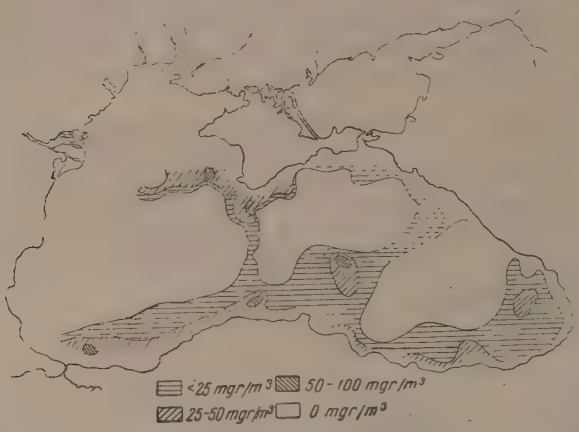

Fig. 2. La distribution horizontale de la biomasse du plancton dans la couche 125—150 m

Le chiffre de la biomasse du plancton subit dans les différentes couches des changements saisonniers. Ces changements atteignent leur maximum dans les couches supérieures (0—25 mg/m³ jusqu'à 203 mg/m³), tandis qu'avec la profondeur ils diminuent graduellement; à 25—50 m la différence entre le minimum et le maximum ne dépasse guère 82 mg; à 75—100 m—45 mg, et dans la zone limitrophe (150—175 m) ces variations sont presque imperceptibles —10 mg/m³ (table).

Changements saisonniers de la biomasse du plancton (mg/m)³

Profondeur, m	V	VI	VII	VIII	IX	X	XI	XII
0—25	120	266	278	335	177	210	187	95
25—50	120	162	202	135	120	148	135	133
75—100	109	80	70	83	84	98	83	64
150—175	39	31	36	38	34	44	40	40

Une série d'agents écologiques, ayant une influence sur les changements de la biomasse du plancton, comme: la température (en premier lieu), la salinité, la lumière, la houle, la quantité de nourriture, subissent des variations considérables dans les couches supérieures de la mer; dans les couches en dessous de 75—100 m ils ne subissent que des variations minimales.

A part cela, tout le qhytoplancton avec ses variations quantitatives et l'apparition périodique de jeunes poissons et de poissons adultes se nourissant de plancton sont, également, concentrés dans les couches supérieures.

En résumant les faits mentionnés ci-dessus, nous pouvons affirmer que le chiffre de la biomasse du plancton subit des variations sensibles dans la couche 0—50 cm et reste, pour ainsi dire, invariable en dessous de 100 m.

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Manuscrit reçu
le 20.VII.1944.

LITTÉRATURE CITÉE

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

PLANT PHYSIOLOGY

VARIETIES OF SUNFLOWER AS CHARACTERIZED ACCORDING TO THEIR DROUGHT RESISTANCE

By T. T. DEMIDENKO

(Communicated by A. A. Richter, Member of the Academy, 12.VII.1944)

The majority of the cultivated plant varieties are distinguished by their high plasticity and are liable to change greatly under external influences. The water regime belongs to the highly effective agents, want or excess of humidity in the soil acting not only upon the anatomical, but also on the internal structure of the plant. When there is a shortage of water in the soil or water is in excess, the plants develop a series of protective adaptations, owing to which they become able better to withstand the drought. During the period of unfavourable conditions the plants develop a xeromorphic structure in their over-ground parts, accumulate more of the «fixed» water, of the hydrophilic colloids, of sugars which increase the osmotic pressure of the cell sap and make the plants more drought-resistant.

A considerable number of varieties created under widely differing climatic conditions are grown both for commercial purpose and in breeding work. It was very important, therefore, to conduct experiments involving a comparative study of the existing varieties of sunflower with regard to their drought resistance. Their drought resistance depends not only on the intrinsic ability of the cytoplasm to endure prolonged withering, but also on the organization of the plant as a whole.

The distribution of the root system is greatly dependent upon the varietal characteristics and the whole habit of the plant, on the structure of the soil and on the depth to which the roots penetrate into the latter. A plant possessing intrinsic properties which make it potentially drought-resistant may nevertheless suffer from the shortage of water if its root system cannot penetrate deep into the soil and remains above the level of the underground water. Another variety possessing less xeromorphic properties may grow and develop normally if it has a wider root system which enables it to absorb the necessary amount of water.

A serious objection to the possibility of estimating drought resistance by vegetative experiments consists in the lack of conditions suitable for a normal growth of the root system. This objection appears doubtful, however, if one takes into consideration the fact that during the droughts of 1936 and 1938 the plants had lost their turgor, which was due to the lack of water in the soil when they were subjected to drought at the period of the formation of the flowering head, of flowering and filling of the seed, although their root system had been fully developed. When drought is prolonged and profound, and there is no sufficient amount of accessible water in the stratum inhabited by the roots, the plants lose turgor and the protective rôle of the root system is no important one. The root system by itself is not an index of drought resistance; thus, varieties possessing similar root systems but differing in the water-retaining ability of the plasm will spend water with different efficiency.

It naturally follows from this that by placing the plants under conditions of scarce water supply, one may experimentally find the reaction of different varieties to the shortage of water in the soil.

The object of our experiment was to test varieties of sunflower for their drought resistance; there were among them both standard varieties and those promising. The experiments were very simple and arranged as follows: the control plants received an optimal amount of water during the whole growing season, while the experimental ones were arranged into three groups, *viz.* those subjected to one drought at the period of the formation of the flower head, to two droughts at the same period plus at the time of flowering, and to three droughts at the time of flower-head formation, of flowering and at the period of seed ripening. These periods were selected because sunflower grown commercially is most often affected by drought at these very developmental stages.

The experiment was conducted in soil cultures, the pots containing six kg of soil each. In order to secure normal nutrition 0.3 g PNK was added per kg of soil. There were four parallel series. The pots were watered twice a day, the humidity being kept at 60 per cent of the full water capacity. The experimental plants were gradually transferred to drought, the quantity of water left in the soil being finally reduced to the «dead store»; this condition was maintained during five days. After this experimental period the water content was at once raised again to sixty per cent to simulate the ordinary conditions existing in commercial growing, when the plants receive water in the form of rain.

The results of yield records are presented in Table 1.

Table 1

Effect of Drought upon Yield of Sunflower

One drought		Two droughts		Three droughts	
Variety	Yield of seed in % to control	Variety	Yield of seed in % to control	Variety	Yield of seed in % to control
3519	108.7	3519	99.74	No.3	51.35
8281	107.89	K638	78.75	No.64	43.30
1846	106.35	1846	73.30	8281	43.11
No.64	103.88	No.3	70.73	3519	38.17
K638	97.37	No.62	64.04	6432	37.94
No.3	96.77	8281	62.23	No.65	37.10
6432	77.48	No.64	62.03	K1483	36.58
K1483	75.18	K1483	54.24	No.62	32.39
No.62	71.09	6432	53.24	K638	32.24
1813	65.20	No.65	53.08	1846	30.56
No.65	64.09	1813	48.60	1813	18.21

It may be observed from the data obtained that different varieties exchange places with regard to their drought-resistance, according to the date of withering. This is caused by their biological nature, *viz.* their different reaction to insufficient water supply at different stages of development.

In the case of a single drought the leading places were held by the varieties 3519, 1846, 8281 and No. 64 which produced higher yield than the control. They were approximated by the varieties K 638 and No. 3 which showed but a slightly reduced yield as compared to the controls.

In the case of 2 droughts the trend was nearly the same, similar varieties taking the first places. Among the varieties subjected to 3 droughts the first place was held by the varieties No. 3, No. 64 and 8281; the varieties K 638 and 1846 appeared in this case to be less drought resistant. Variety 1813 proved inferior in all the three experiments. Variety No. 65 after one and two droughts came last or last but one, while in the three-drought experiment it wandered to the sixth place; this latter fact shows that the period of seed ripening is the

most critical one in this variety. The remaining varieties took up intermediate positions.

Hydrophilic colloids play an important rôle in the protective properties of the tissues against withering and drought; their amount varies during the growing period in each variety with growth and development, as well as with the number of droughts experienced by the plant during the summer.

We shall present here only a part of the results of the numerous determinations of colloids made by us in sunflower (Table 2).

Table 2

Accumulation of Colloids in Leaves of Sunflower Repeatedly Subjected to Drought

Variety		Colloids, per cent		
		Total	Reversible	Irreversible
3519	Control	6.96	6.56	0.50
	One drought	8.08	7.88	0.20
	Two droughts	8.32	7.86	0.48
	Three »	9.38	8.35	1.03
8281	Control	6.83	6.50	0.33
	One drought	7.88	7.50	0.38
	Two droughts	7.12	6.84	0.26
	Three »	8.92	8.26	0.66
1483	Control	8.02	7.76	0.26
	One drought	8.94	8.74	0.20
	Two droughts	8.72	8.39	0.33
	Three »	8.84	7.62	1.24
1813	Control	9.16	9.09	0.07
	One drought	8.62	7.73	0.36
	Two droughts	8.48	8.13	0.36
	Three »	8.38	8.06	0.32

It may be observed from Table 2 that the amount of reversible colloids accumulated in the leaves of sunflower exceeds that of irreversible ones.

One may see a difference in the accumulation of colloids under the influence of withering in drought-resistant and non-resistant varieties. While in the varieties 3519 and 8281 the amount of reversible and irreversible colloids was increased under the influence of drought, in the variety 1813 it diminished. The amount of colloids in each experiment has changed but slightly in the intermediate variety K 1483 with the increased number of witherings.

Table 3 is intended to give an idea of in how far do the experimental results depict the real drought resistance of the varieties of sunflower under field conditions.

Table 3

Effect of Drought upon Yield of Sunflower Varieties
(data from variety test and from the experiments)

Data from variety test		Vegetation experiments of 1939					
		One drought		Two droughts		Three droughts	
Variety	Yield of seed in 1938 in % to 1939	Variety	Yield of seed in 1938 in % to control	Variety	Yield of seed in 1938 in % to control	Variety	Yield of seed in 1938 in % to control
3519	90.54	3519	108.7	3519	99.74	No.64	43.30
1846	85.06	8281	107.89	1846	73.30	8281	43.11
No.64	84.96	1846	106.35	No.62	64.04	3519	38.17
8281	83.23	No.64	103.88	8281	62.23	6432	37.94
6431	81.84	6432	77.48	No.64	62.03	No.65	37.10
No.62	80.65	No.62	71.09	6432	53.24	No.62	32.39
No.65	77.94	1813	65.20	No.65	53.08	1846	30.56
1813	77.87	No.65	64.09	1813	48.60	1813	18.21

Table 3 contains the results of a variety test, obtained in 1938 and 1939, the two yields greatly differing with regard to humidity: in 1939 there was a great amount of atmospheric precipitation, while the year 1938 was distinguished by a great drought. The results of the field observations and of the experiment coincide on the whole. Thus, variety 3519 held the first place with regard to drought resistance both under field conditions and in experiments with one and two droughts. Good coincidence is also found in the variety 1813.

Certain discrepancy in a number of varieties may be explained by the great influence which is exerted upon drought resistance by the differently developed root system.

Conclusions. 1. Vegetation method is entirely suitable to the comparative variety tests for drought resistance.

2. In drought-resistant varieties the amount of hydrophilic colloids (reversible and irreversible) is augmented under the influence of withering, while in non-resistant ones the reverse phenomenon is observed.

The Krasnodar
Plant-Breeding Station.

Received
6.VII.1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

PLANT PHYSIOLOGY

A NEW METHOD FOR GRAFTING HERBACEOUS PLANTS, APPLICABLE UNDER FIELD CONDITIONS

By M. PARKHOMENKO

(Communicated by A. A. Richter, Member of the Academy, 12. VIII.1944)

Grafting was known in gardening and viticulture, both of which deal with perennial woody plants, as early as over two thousand years ago.

Vegetable gardening and floriculture which, on the contrary, deal mainly with annual herbaceous plants, do not use grafting. The need in grafting of herbaceous plants, however, becomes more and more imperative, not only on economic grounds, but also in connexion with the fact that annual herbaceous plants represent material highly suitable for the study of the relations existing between stock and scion. Owing to their short growing period, it is possible to obtain several generations within a year.

Grafting of herbaceous plants is not very successful, however. Even under laboratory conditions where various precautions can be taken (covering with slips, darkening, low temperature) the proportion of successful grafts does not exceed 10 or 15 per cent. In the field this proportion drops down to 1—2 per cent.

None of the generally known methods can ensure higher proportion of success; all these methods were worked out for trees and automatically applied to herbaceous plants. We have worked out a simple method which is not laborious, easily accessible and requiring no precautions; when applied to herbaceous plants, it gives hundred per cent success not only at the laboratory, but in the field, too. Our method was checked up in the course of three year's work on over a thousand graftings of different herbaceous plants. It gave uniformly positive results both in the field and at the laboratory. We therefore find it possible to recommend it as a wholly dependable method for grafting herbaceous plants, applicable under commercial field conditions.

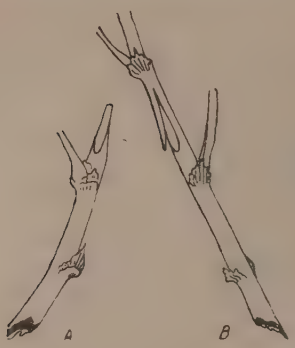
Plants intended for grafting are grown together in the nearest proximity to each other. The distance between the prospective stocks and scions depends either on the height or length, as well as on the elasticity of the stems or runners of the plants to be grafted. In any case such a distance should be selected that the stock and scion can easily be brought into contact at the point when grafting is to be done, remaining on the roots of their own. Two examples would serve as illustration. 1) Sunflower and girasole (*Helianthus tuberosis*). The two plants possess rigid stems. In this case the distance between the stock and scion should not exceed 8—10 cm. 2) Water-melon and melon. Both form long elastic runners. The distance between the prospective scion and stock may be brought to one metre and even more.

The dates of the planting of the stock and scion may coincide or may not. They are chosen on purely technical grounds, since different plants form shoots suitable for grafting with different speed, or on economical or scientific grounds. The most

common case would probably be the grafting of plants of equal age. It may occur, however, that fully formed stock and young scion (or *vice versa*) would be required. As a matter of course, the dates of planting would be different in this case. Anyhow the time of planting should be chosen in such a way as to make the stock and scion develop technically suitable material of a corresponding age.

When scion and stock are ready to be grafted, one has to determine in a preliminary way the points at which the two should be grafted together. The very point of grafting is determined on various grounds which will be discussed below.

After the point when grafting is to be done is selected, the apex of the stock is cut clear away at this point. The detached segment is not needed for the future work, and discarded. Then the stock is shaped wedge-like to form a wedge 2—4—6 cm long and an angle of 5—7° (figure, *A*). This done, a longitudinal cut is done upwards through the stem of the scion in such a manner as to bring in touch the inner end of the cut and the core of the shoot. The length of the cut should be equal to the length of the wedge of the stock or even somewhat exceed it (figure, *B*).



The wedge of the stocks should be cautiously and without delay inserted into the cut made on the scion. The parts of the graft are carefully arranged in proper order, and a bandage is made to cover the wounded place. The first step is made.

The material distinction of our method, as shown above, consists in the fact that the scion does not lose its connexion with the root system of its own at the moment it is transplanted to an alien root system. At the moment of grafting only half of its conducting system is damaged. This is why, even in the open, in full sun such grafts not only escape death, but do even not wither noticeably. Growing together takes place in a considerably shorter time than with all other methods of grafting. There are grounds to believe that with interfamiliary grafting this method will yield very satisfactory results.

As soon as the graft has taken (after 6—10 days), the bandage should be removed. In case the bandage is left any longer, it would cut into the thickening stem, which may damage the graft and interfere with the removal of the bandage. After the bandage is removed, there may arise the danger of breakage of the graft-seam. Whenever it may seem necessary, one should cover the graft again with two or three turns of the bandage.

Finally, when the graft has fully set, it remains to make a cut clear across that part of the combinations which leads to the roots of the scion; the cut should be made somewhat obliquely, downwards. From this moment on the combination will acquire the appearance of an ordinary graft. The technical part of the operation is thus completed.

Our method can afford the production of more complex graft combinations which may be of special interest both scientifically and commercially. The following examples may illustrate the idea.

After the primary bandage is removed and the secondary one put on, a graft combination is obtained where the scion receives water and mineral food from two distinct root systems, the foreign and its own. Such a combination by itself is of interest to the physiologist and the genetist. The new method of grafting will make it possible to study the behaviour of the scion transferred to nutrition from two root systems; what would be the behaviour of these root systems possessing a common foliage? One will now be able to study such combinations where the two root systems belong to a common species, or they originate from different species, genera or even families. Finally, the scion may be grafted to three or more root systems.

Let us take another example. Let us graft the melon to the so-called hundred-pound pumpkin. After the graft had grown together, the root system of the melon

is separated from the graft. The grafting should be done in such a way as to make a female flower of pumpkin be formed immediately beneath the point of grafting. The flower should be pollinated with the pollen of the hundred-pound pumpkin. Then all the leaves and side shoots are carefully removed from the pumpkin-plant, while all the foliage is preserved on the melon-plant. In such a way the developing fruit of the pumpkin, as well as its seeds are brought up on the assimilates of the scion, the melon.

The combination may be made more complex still. The graft between the table water-melon and the fodder water-melon may serve as an example. Grafting is done in such a way as to obtain a female flower of the scion, the table water-melon, just above the point of grafting. After grafting had taken, the flower is pollinated with the pollen of the table water-melon. Then the shoot of the table water-melon is cut off, a little above the flower; the assimilation system of the scion is thus removed entirely. Conversely, all the leaves and side-shoots of the fodder water-melon are preserved and kept in good condition. The fruit of the table water-melon and, naturally, the embryos developing in it, would be formed only at the expense of the foreign substances coming from the root system and assimilation tissue of the fodder water-melon.

It is clear from the above examples that the place and type of grafting may vary considerably, according to the aim of experimenter.

Our aim was to work out a method for grafting herbaceous plants which would yield a high proportion of taking and suitable to all circumstances, including field conditions. At the same time we intended to show that such a method opens wide prospects for the application of grafts of herbaceous plants to a variety of problems, both theoretical and purely practical.

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Ashkhabad.

Received
3.VI.1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

PHYTOPATHOLOGY

ON A DEFENSIVE REACTION OF THE PLANT CELL

(A CONTRIBUTION TO THE COMPARATIVE PHARMACOLOGY OF SULPHONAMIDES)

By V. L. RISCHKOV

(Communicated by A. A. Richter, Member of the Academy, 28.X. 1944)

The sulphonamides which have acquired so great an importance in medicine are of interest also in a broader, more general biological aspect on account of the relation which has been found to exist between the ability of these compounds to depress the growth of cells and the amount of *p*-aminobenzoic acid contained in the latter. There is nothing surprising therefore in the fact that numerous investigations have been made public, dealing with the influence of sulphonamides not only on bacteria and higher animals but on plants as well.

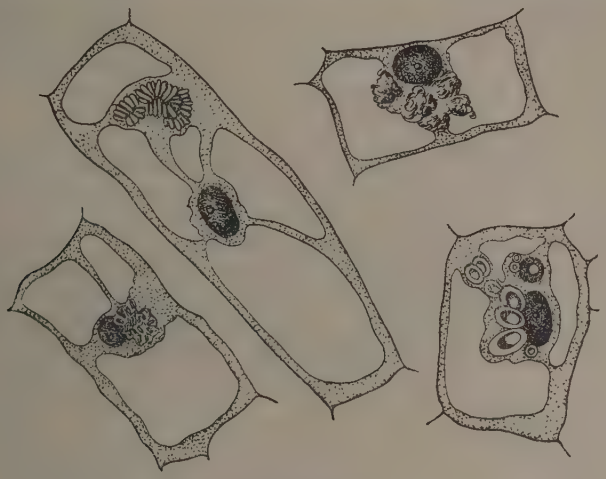
The same causes induced us to investigate the question of the effect of sulphonamides on the growth of higher plants. In the course of this study we found that red soluble streptocide may sedimentate within the plant cell in the form of crystals; this circumstance, having been subjected to a more detailed study, became the starting point for a number of comparative considerations of a more general character, which we report in the present article.

Tomato seeds were germinated on filter paper soaked in a 1—0.5 per cent solution of red soluble streptocide (sodium 4-sulphonamido-phenylazo-1-acetylamino-8-oxy-3,6-naphthalene-disulphonate). Under these conditions the energy of growth of the seeds differs but little from the one they display on filter paper soaked in water. The developing roots assume a deep red colour. This coloration does not disappear if the seedlings are transferred into water, but the lateral rootlets developing later and the newly formed tissues of the tap-root are no more stained by streptocide, which shows that this substance, after it has been absorbed by the tissues, does not spread into new cells which are formed later, to any significant degree. If such a seedling with its root stained with streptocide is planted into soil, it can live and develop there for a long time, and the coloured tissues keep the stain for months. Having made sure of the property of streptocide to stain durably the tissues of the plant, we turned to the investigation of the conditions in which this substance keeps within the cells.

The investigation was conducted on cross-sections through the root prepared with the aid of a razor so that the cells could be studied *in vivo*. It was found that streptocide is, as a rule, accumulated in the cells of the tomato root in an insoluble state, forming within their protoplasm accumulations of red crystals. At first it was supposed that streptocide suffered some chemical transformation within the cells and became insoluble as a result of this transformation. It was found, however, that it suffices to subject the plant cell to some mechanical injury or to kill it in any other way, in order that the crystals dissolve immediately, the cell, including its envelope, staining red throughout. It was thus found that the ability

to accumulate streptocide in the form of crystals is a function of the live cell. This seemed particularly interesting, since a similar phenomenon is known in regard to the thoroughly investigated crystals of the tobacco mosaic virus in the cells of mosaic stricken plants (the Iwanowsky crystals). Every treatment leading to the death of the cell leads to the dissolution of these crystals ([1]).

This induced us to acquaint ourselves in greater detail with the mechanism of the accumulation of masses showing a granular or spongy structure, inside of which small crystals of streptocide begin to form. Later on these structures are entirely filled up with large crystals, and the stroma on the base of which they have arisen becomes almost indiscernible (figure). It should be reminded that the crystals of the virus of tobacco are also accumulated in the plant cell on the basis of a change originating initially in the protoplasm itself. Many authors have observed the formation of the crystals of virus protein inside of initially amorphous amoeboid



Cells of tomato root with accumulations of red soluble streptocide. Accumulations of protoplasmatic bodies in which crystals of streptocide are formed, vacuolated bodies and bodies in the shape of discs stained with streptocide are visible in the cells

corpuscles—a process which outwardly resembles greatly the early stages of the formation of streptocide crystals within amorphous bodies.

The analogies drawn up above permit to advance the hypothesis that both mechanisms, that of the formation of virus crystals and that of the formation of the crystals of streptocide, are closely related and represent a defensive reaction on the part of the cell. This reaction consists in the arising of differentiations, which are probably coacervates, similar to those described of late by Dufrénoy ([5]), within the protoplasm under the influence of the foreign substance penetrating into the cell. It is in these structures that the foreign substance is accumulated, a relative blockade of the substance being achieved by this means. At the same time the foreign substance, owing to its concentration in the coacervates, may crystallize there.

The structures in which soluble streptocide is blocked in the cells of the tomato do not always contain it in crystalline form. In numerous cells we found great numbers of corpuscles stained red by streptocide. Sometimes these corpuscles were oval and had one or several vacuoles, in other cases they resembled a capsule with an aperture, in others again they were the shape of a ring, in some more they resembled small discs flattened in the middle.

Everyone who has studied pathological cytology of plants is well acquainted with bodies of this kind. In virus diseases they have been described on many occasions as amoeboid or vacuolated bodies. Structures of this type have been

found to exist in potato tubers by the present author ([2]). In this case they proved to be tannoid bodies, and it was established that these very tannoid bodies had been assumed by Rozé ([6]) to be none other than the ubiquitous parasite *Pseudocommunis*. The investigation of these bodies in the tubers of potatoes and their comparison with the data reported by Rozé induced already at that time to think that corpuscles of this type arise in the cell under the influence of various noxious agents. We asked therefore Sukhov and Lanshina to make an attempt at inducing their formation by treating beet seedlings with potassium iodide. This was achieved. Dufrénoy considers such structures as coacervates of tannoids, phospholipids and other substances with the proteins of the protoplasm.

As it has already been said, the coacervates described here were stained red by streptocide, which was absent from the rest of the protoplasm and the vacuoles, at least in quantities discernible under the microscope. This shows that inside of the cell a substance foreign to it may be blocked in such coacervates.

The reported observations show that the formation of different coacervates under the influence of various agents within the plant cell has the character of a defensive reaction. The coacervates appear to be centres of the concentration of foreign substances which either penetrate into the cell from outside, or are formed within the cell itself under the influence of some pathological process. It is hardly possible to doubt that the accumulation of virus crystals, as well as the formation of the various amoeboid or vacuolated bodies in virus diseases, are only particular cases of the manifestation of the defensive reaction of the cell, described in the present article.

The special interest of these findings for virology is that they help us to understand the conditions of the accumulation of the hexagonal crystals of virus protein, found in the cells of a mosaic stricken tobacco plant. The mechanism of the formation of crystals and paracrystals of an entirely different type which has been dealt with in an earlier study by Rischkov and Smirnova ([3]), is not considered here.

Moscow Medical Institute.
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Received
28.X.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

EXPERIMENTAL MORPHOLOGY

TIME OF DETERMINATION OF A TUMOUR IN THE PROCESS OF ITS INDUCTION WITH 3:4 BENZPYRENE

By N. I. LASAREV

(Communicated by I. I. Schmalhausen, Member of the Academy, 8. VII. 1944)

By introducing the conception of «determination» into oncology we do not merely mechanically adopt this term from experimental embryology. Both in older literature and in the more recent one we meet with attempts to draw analogies between the processes of tumour development and organogenesis. Indeed, the existence of «common» features in a number of the details of both processes has been established by the latest researches in the field of chemical embryology. Alongside of the detection in tumour cells of such characters as are peculiar to embryonic cells, the experimental induction of «teratome-like» formations from the cells of embryos, developing out of delayed frog eggs (formations which, being grafted to tadpoles, produce typical tumour growth with metastasis) ([1]), alongside of the discovery of endogenic blastomogenic substances in the organism ([2]), the development of a complementary nervous system has been obtained in the embryo by acting upon it with blastomogenic hydrocarbons ([3]). There can be no doubt that the last fact points to a certain «community» of the two processes induced by the same agent. The difference in the effect produced by blastomogenic substances in this case cannot be explained by anything else than the difference between the reactive capacity of embryonic tissue and the tissue of the adult organism(morphogenetic potency).

The introduction of the conception of the «determination» of the tumour rudiment is justified both from the clinical point of view and from that of experimental oncology. The terms «formative irritation» and «evocator», as applied to the understanding of the mechanism of action of blastomogenic agent, have not been taken at random either ([4]). Borst ([5]) has assumed in apparently normal malignant cells the existence of certain pathological characters lacking morphological expression. Fischer-Wasels ([6]) had advanced the conception of the «tumour rudiment», and modern oncology has accepted the conception of «pre-cancer», understanding by this term a condition predetermined to a certain degree and arising long before the manifestation of the morphological symptoms of the tumour. Lastly, Shear ([7]) succeeded in 1936 in producing a tumour at the site of the graft of a part of the connective-tissue capsule, which had formed around cancerogenic substance on the 47th day after the inoculation of the latter, *i. e.* at such a date when there were as yet no morphologic signs of tumour growth in the material used for the graft. Thus, by the present time certain data are already available which permit to suppose that the determination of the tumour rudiment precedes its morphological manifestation.

On the basis of the data available , we undertook to arrange an experiment for the investigation of one of the most elementary problems of the determination of

the tumour rudiment, that of the time of its determination. The technique of this experiment had to provide for the possibility of extracting the blastomogenic substance after a definite interval of time from the beginning of its action. The most suitable method for this aim is to introduce the substance in the shape of paraffin pills. 3 : 4 benzpyrene was introduced into liquefied paraffin so that each pill should contain 3 mg of this compound. Such a pill was placed into a glass tube, one end of which was cut off slantwise like a syringe needle. An incision was made in the skin of a mouse near its tail, the tube was inserted through it under the skin and made to reach the interscapular region. There the pill was pushed out of the tube with the help of a glass rod. The site between the shoulder plates was chosen to prevent the mouse from gnawing the pill.

According to Roussy ([8]) the interscapular region is most susceptible to the action of blastomogenic agents. It was found in the course of the very first days that in an overwhelming majority of cases the pill fell out spontaneously owing to necrosis and perforation of the skin, and in the following a bit of thin glass (a cover-slip) was placed between the pill and the skin to prevent the lesion of the latter.

A total of 127 mice were under experiment. Every 3—4 days the pills were extracted from 2—3 mice by means of an incision through the skin and the connective-tissue capsule, which was evident already a few days after the inoculation of the pill. Part of the mice died, another part were used for supplementary investigations, so that the main experimental series included 72 mice only.

Tumours appeared in 21 animals; in 9 of them there were multiple tumours in sites distant from the site of inoculation (mammary glands and lungs); in 12 mice various forms of cancer (2 cases) and sarcoma (10 cases) developed at the site of the localisation of the pill. All the tumours appeared after a relatively short period, 150 to 210 days from the moment of the inoculation of the pill, and in those mice in which no tumours developed during this period, they failed to appear later either.

In the course of the subsequent study of the data thus obtained the following phenomenon stood out particularly. All the 9 animals in which distant tumours had developed belonged to the group in which the excision of the pill was done between the 22nd and the 30th day from the moment of its inoculation, no tumour developing at the site of the pill in any of them. All the 12 mice in which tumours had developed at the site of the inoculation of the pill, belonged to those in which the excision was made after the pill had remained more than 30 days in the organism of the animal, no tumours developing in them at any other site. The mice in which no tumours whatever were registered may also be divided into two groups: the first one includes 23 mice in which the pill was excised during the first 22 days of its stay in the organism, the second—20 mice in which it was excised alter the 22nd day. No tumour appeared in any of the mice belonging to the first group, and the second group is characterized by the non-appearance of any tumour, while it had appeared in other mice subjected to the action of the pill for the same period of time.

This strict dependence of the results of our experiments on the time of the action of the pill suggested the idea that it is possible to divide the process of tumour development into three stages.

The first stage embraces 30 days of the action of the pill. This is the minimum time, during which the action of the blastomogenic substance is absolutely necessary to induce the development of a local tumour.

It is in the course of this time interval that the «tumour rudiment» arises, or, in other words, the determination of the tumour takes place. This stage, accordingly, should be designated as the stage of determination. It equalizes only about $^1/_5$ of the time necessary for the arising of a morphologically apparent tumour. It is perfectly evident that 30 days is only the minimum time of action necessary for a tumour to be induced, but it is not clear as yet,whether this time of action is also the optimal time of tumour determination.

The stage of determination in its turn may be subdivided into two periods. The first of these, comprising 22 days, represents the minimum time necessary for

inducing distant tumours; this interval of time is, however, insufficient for the determination of a local tumour, for which a greater time interval is necessary. The production of distant tumours by a shorter exposure to the action of the pill containing a blastomogenic substance conforms well with the results obtained with the aid of reduced doses ([9]) of the blastomogenic agent, or with that of slightly cancerogenic slate tars ([10], [11]).

The second stage is the latent period which, as this has been demonstrated by the study of 50—60-day-old capsules under tissue culture conditions, made by Zorin, the capsule may acquire certain biological tumour characteristics. Thus, in spite of the absence of any morphological tumour symptoms, the culture of capsules revealed such peculiarities of growth as are typical of the growth of sarcoma tissue. It should be emphasized that a latent period is a natural character of every organogenesis.

The beginning of the third stage—the stage of the expression of the tumour—starts from the moment of the manifestation of the acquired properties, when the tumour can already be recognized morphologically. The beginning of this stage is evidently the end of the process of development of the tumour.

Summing up the analysis of the obtained results, the following conclusions may be drawn:

1. In the course of the process of development of the tumour a determination of its rudiment takes place, similar to the rudiment determination in organogenesis.

2. The time of determination of the tumour rudiment, according to the results of our experiments, is equal to 30 days of exposure to the action of the 3 : 4 benzpyrene. This time is evidently corresponding to the beginning of the «pre-cancer condition».

3. Distant tumours, the appearance of which was observed, according to literary data, after exposure to smaller doses or after treatment with weaker blastomogenic agents, may be induced also by shortening the time of exposure.

4. The process of the development of the tumour should be divided into three separate stages: 1) the stage of determination of the tumour rudiment, 2) the latent period of the process of development of the tumour, and 3) the stage of the expression of the tumour.

Received
8.VII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 7

HELMINTHOLOGY

ON THE TAXONOMIC POSITION OF THE GENERA *ACANTHOCHEILONEMA* COBBOLD AND *MOLINEMA* FREITAS ET LENT WITHIN THE SYSTEM OF NEMATODES

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The genus *Dipetalonema* was founded in 1861 by Diesing (²) for a nematode described in 1858 by Molin (¹) from the intermuscular tissue of various Brazil monkeys, under the name of *Filaria caudispina*. Linstow (1899) made a suggestion as to the identity of this species with *Filaria gracile* Rud., 1809; as a result, the latter species was, in conformance with the rules of nomenclature, legalized by Yorke and Maplestone (1926) as a type-species of the genus *Dipetalonema* (⁵). *D. gracile* was thoroughly investigated by these two authors, who found it to bear on the cephalic end a peribucal cuticular shield, rectangular in shape, laterally elongated and supplied with two lateral and four submedian papillae. This rather essential morphological character has rightly been introduced by Yorke and Maplestone into the diagnosis of the genus *Dipetalonema*, which enabled them to include the said genus into the subfamily *Setariinae* which had newly been founded by these authors.

The diagnosis of the genus *Dipetalonema*, as formulated by Yorke and Maplestone, implies also the presence upon the caudal end of both males and females of two conic processes, owing to which the tail assumes a trifurcate configuration, as duly described and illustrated in Molin's (1858) diagnosis of *Filaria caudispina*, and also alluded to by the specific name of the latter.

In 1870 Cobbold described *Filariae* from African dogs and hyaenas, and founded a new genus and species, *Acanthocheilonema dracunculoides*.

As may be seen from the description made by Cobbold himself (³), as well as from the subsequent papers by Railliet, Henry and Langeron (⁴), this nematode has a simple cephalic end, lacking any traces of a peribucal cuticular shield; while the caudal end in both male and female bears a pair of lateral conical annexes similar to those in *Dipetalonema gracile*.

Until 1926, both *Dipetalonema* and *Acanthocheilonema* were referred to in literature as two independent genera, either of which comprised at that time several independent species, described by various authors.

In 1926 the two genera were united into a single genus by Yorke and Maplestone (⁵); as a result, the genus *Acanthocheilonema* has become the synonym of the genus *Dipetalonema*. These authors referred to the genus *Dipetalonema* ten species, having included into it, alongside of unquestionable representatives of this genus, all the species of *Acanthocheilonema*, inclusive of the type species founded by Cobbold, *viz. A. dracunculoides*.

The genus *Dipetalonema* was thus erroneously made to comprise nematodes with similar caudal ends (trifurcate), but showing entirely different structures of the ornamentation of their cephalic end.

As a matter of regret, Yorke and Maplestone have never become aware of their mistake; it was also overlooked by other writers who recognized *Acanthocheilonema* as a synonym of *Dipetalonema*. Even Brumpt (1936), though referring in the 5th edition of his «Précis de parasitologie» the genus of *Filaria* parasitic of man to the genus *Acanthocheilonema*, makes in a foot-note to page 979 a restricting remark that the priority belonged to the genus *Dipetalonema* Diesing, 1861.

Faust (1929) withheld himself from expressing his personal opinion as to the identity of the two genera; yet, the species *Filaria perstans* Manson, 1891, has been referred by him to the genus *Acanthocheilonema*, instead of *Dipetalonema*. And lastly Walton (1927), in his description of two *Acanthocheilonema* species, remarks in a rather elusive way that if Yorke and Maplestone's viewpoint be adopted, the species in question should, as a matter of fact, be referred to *Dipetalonema*.

In 1939 Freitas and Lent (6) founded a new genus, *Molinema* (subfamily *Setariinae*) for two species described by Melin, *viz. Filaria diacanthum* (included by Yorke and Maplestone into the genus *Dipetalonema*), and *Filaria bifida*. Although the caudal annexes in these species are conical in shape, they are entirely deprived of a peribucal shield, which makes it judicious to eliminate *F. diacanthum* from the genus *Dipetalonema*.

Now the fact that the *Molinema* species are deprived of any ornamentation in the region of the cephalic end, testifies to this genus being erroneously included into the subfamily *Setariinae*.

In the course of our work on a monograph devoted to the family *Filariidae*, we made an analysis of all the controversial data in which abounds the literature concerned with this group of nematodes; the conclusions to which we have been led by this work are the following:

1. The genus *Dipetalonema* Dies., 1861, must embrace such species only as display a correlative combination of a cuticular peribucal shield and a trifurcation of the caudal end. Such species are: *D. gracile* (Rud., 1809), *D. evansi* (Lewis, 1882) and *D. multipapillata* Walton, 1927.

The question whether or not *D. roemeri* Linstow, 1905, belongs to this genus, cannot be answered until subjected to a more scrupulous analysis.

2. The genus *Acanthocheilonema* Cobbold, 1870, should be rehabilitated as an independent genus. It must not be considered synonymous with *Dipetalonema*, in so far as the species it comprises are all deprived of a peribucal cuticular shield. The following species are to be referred to it: *A. dracunculoides* Cobbold, 1870; *A. perstans* (Manson, 1891); *A. grassii* (Noé, 1907); *A. reconditum* (Grassi, 1890); *A. weissi* Seurat, 1914.

Aside from these species, we refer to the genus *Acanthocheilonema* the species *Filaria hippopotami* Leiper, 1910, from the hippopotamus.

Whether it is possible to refer to this genus also *Filaria australe* Linstow, 1897, the males of which bear no trifurcation on the apex of their tails, is a question requiring closer investigation.

3. The representatives of the genus *Acanthocheilonema* Cobbold, 1870, and *Molinema* Freitas et Lent, 1939, are entirely lacking any cuticular ornamentation of their cephalic ends, on account of which they are to be excluded from the subfamily *Setariinae*, and referred to the subfamily *Filariinae*. The genus *Dipetalonema* in its new, restricted form must remain within the subfamily *Setariinae*.

4. The diagnosis of the genus *Acanthocheilonema* is formulated by us in the following way: *Filariinae* with a thread-like, delicate body covered with a cuticle either smooth or ornamented with longitudinal, occasionally transverse lines.

Mouth simple, without labiae or oral ornamentation. Oesophage distinctly subdivided into two parts. Caudal end both in female and male bears at its apex two laterally arranged short conic annexes, owing to which the end of the tail assumes a trifurcational outline. These annexes differ in their expression in different species.

An exclusion to the general rule is offered by the males of *A. australe* which, according to Linstow, have no annexes upon the caudal end of their bodies. M a l e: Caudal end coiled in the form of a spiral. Caudal wings absent. As a rule, four

pairs of pre-anal papillae present. The number of post-anal papillae variable. Two different spiculae: the larger, membranous in its distal part; the smaller, usually bearing a minute hooklet at its end. Gubernaculum lacking. Female: Vulva situated in the region of the oesophage. Viviparous. Pass larval stage in the blood of carnivora, primates and rodentia, being parasitic of the serous tissues and subcutaneous fat tissue of these animals.

Type species: *A. dracunculoides* Cobbold, 1870, from dogs and African carnivora.

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Received
3.II.1945.

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